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MATHÉMATIQUES

SUR LES INTÉGRALES DES ÉQUATIONS AUX DÉRIVÉES PARTIELLES DU PREMIER ORDRE D'UNE FONCTION INCONNUE, QUI NE CONTIENNENT PAS LA FONCTION INCONNUE ET SONT HOMOGÈNES PAR RAPPORT AUX DÉRIVÉES

Par G. PFEIFFER, de l'Académie des Sciences de l'Ukraine

Dans le présent article nous traitons une question, que l'on ne considère pas dans les manuels fondamentaux et à laquelle des auteurs de nombreux travaux scientifiques prêtent peu d'attention.

L'équation aux dérivées partielles du premier ordre d'une fonction inconnue

$$f\left(x_1, x_2, \ldots x_n, \frac{p_1}{p_n}, \frac{p_2}{p_n}, \ldots \frac{p_{n-1}}{p_n}\right) = 0 \tag{1}$$

possède les intégrales complètes de Lagrange de trois formes:

A) L'intégrale sans facteur constant arbitraire et sans constante arbitraire additive

$$\left.\begin{aligned} &z = \Phi(x_1, \ldots x_n, a_1, \ldots a_n),\ J = \frac{D(\Phi_1, \Phi_2, \ldots \Phi_n)}{D(a_1, a_2, \ldots a_n)} \equiv 0, \\ &J_k = \frac{D(\Phi, \Phi_1, \ldots \Phi_{k-1}, \Phi_{k+1}, \ldots \Phi_n)}{D(a_1, a_2, \ldots \ldots \ldots \ldots a_n)} \not\equiv 0 \ (k - \text{l'un des nombres } 1, 2, \ldots n), \\ &J_{kj} = \frac{D(\Phi_1, \ldots \Phi_{k-1}, \Phi_{k+1}, \ldots \Phi_n)}{D(a_1, \ldots a_{j-1}, a_{j+1}, \ldots a_n)} \not\equiv 0 \quad (j - \text{l'un des nombres } 1, 2, \ldots n), \\ &R_\tau^n = \frac{D\left(\frac{\Phi_1}{\Phi_n}, \frac{\Phi_2}{\Phi_n}, \ldots \frac{\Phi_{n-1}}{\Phi_n}\right)}{D(a_1, \ldots a_{\tau-1}, a_{\tau+1}, \ldots a_n)} \equiv 0 \ (\text{pour tous les } \tau = 1, 2, \ldots, n). \end{aligned}\right\} \tag{2}$$

B) L'intégrale sans facteur constant arbitraire, ayant une constante arbitraire additive

$$\left.\begin{aligned} &z = \Theta(x_1, \ldots x_n, a_1 \ldots a_{n-1}) + a_n,\ L_k = \frac{D(\Theta_1, \ldots \Theta_{k-1}, \Theta_{k+1}, \ldots \Theta_n)}{D(a_1, a_2, \ldots \ldots \ldots a_{n-1})} \not\equiv 0, \\ &\qquad (k - \text{l'un des nombres } 1, 2, \ldots n); \\ &L = \frac{D\left(\frac{\Theta_1}{\Theta_n}, \frac{\Theta_2}{\Theta_n}, \ldots \frac{\Theta_{n-1}}{\Theta_n}\right)}{D(a_1, a_2, \ldots \ldots \ldots a_{n-1})} \equiv 0. \end{aligned}\right\} \tag{3}$$

C) L'intégrale à facteur constant arbitraire, ayant la constante arbitraire additive, l'intégrale de A. Mayer (1):

$$\left.\begin{aligned} &z = a_{n-1}\varphi(x_1, \ldots x_n, a_1 \ldots a_{n-2}) + a_n, \\ &N_k = \frac{D\left(\frac{\varphi_1}{\varphi_n}, \ldots \frac{\varphi_{k-1}}{\varphi_n}, \frac{\varphi_{k+1}}{\varphi_n}, \ldots \frac{\varphi_{n-1}}{\varphi_n}\right)}{D(a_1, a_2, \ldots \ldots \ldots \ldots a_{n-2})} \not\equiv 0, \\ &(k - \text{l'un des nombres } 1, 2, \ldots n-1). \end{aligned}\right\} \tag{4}$$

D'après l'intégrale A)

$$z = \Phi(x_1, \ldots x_n, a_1, \ldots a_j, \ldots a_n) \tag{5}$$

il est facile de construire l'intégrale B)

$$z = \Theta(x_1, \ldots x_n, a_1, \ldots a_{j-1}, a_{j+1}, \ldots a_n) + \text{const}; \tag{6}$$

il ne faut que remplacer dans l'intégrale (5) la constante arbitraire a_j par une quantité fixe, arbitrairement prise, d et d'ajouter à la fonction Φ une constante arbitraire:

$$\begin{aligned} z &= \Phi(x_1, \ldots x_n, a_1, \ldots a_{j-1}, d, a_{j+1}, \ldots a_n) + \text{const} = \\ &= \Theta(x_1, \ldots x_n, a_1, \ldots a_{j-1}, a_{j+1}, \ldots a_n) + \text{const}. \end{aligned} \tag{7}$$

Les conditions, liées à l'intégrale (5), deviendront des conditions pour l'intégrale (6) [2].

Dans l'intégrale B)

$$z = \Theta(x_1, \ldots x_n, a_1, \ldots a_i, \ldots a_{n-1}) + a_n \tag{8}$$

il existe toujours une constante arbitraire a_i (et peut être plusieurs), qui remplacée par une quantité fixe, arbitrairement prise, d permet de former l'intégrale C):

$$z = a_0 \Theta(x_1, \ldots x_n, a_1, \ldots a_{i-1}, d, a_{i+1}, \ldots a_{n-1}) + a_n \text{ [3]}. \tag{9}$$

À l'intégrale C)

$$z = a_{n-1} \varphi(x_1, \ldots x_n, a_1, \ldots a_{n-2}) + a_n \tag{10}$$

correspond l'intégrale de S. Lie de la première classe

$$z = \text{const}, \quad \varphi = \text{const} \text{ [4]}. \tag{11}$$

D'après l'intégrale (11) ou, ce qui est le même, d'après l'intégrale (10), nous construisons immédiatement les intégrales [5]:

$$\text{A)} \quad z = \Phi(\varphi, a_{n-1}, a_n), \tag{12}$$

où Φ est une fonction fixe, arbitrairement prise, de ses arguments, qui ne contient pas de constante arbitraire additive et satisfait à l'une des conditions

$$\frac{\partial^2 \Phi}{\partial \varphi \partial a_n} \not\equiv 0, \quad \frac{d}{d\varphi}\left(\frac{\frac{\partial \Phi}{\partial a_n}}{\frac{\partial \Phi}{\partial a_{n-1}}}\right) \not\equiv 0, \tag{13}$$

$$\frac{\partial^2 \Phi}{\partial \varphi \partial a_{n-1}} \not\equiv 0, \quad \frac{d}{d\varphi}\left(\frac{\frac{\partial \Phi}{\partial a_n}}{\frac{\partial \Phi}{\partial a_{n-1}}}\right) \not\equiv 0 \tag{14}$$

ou bien à toutes les deux.

$$\text{B)} \quad z = \Theta(\varphi, a_{n-1}) + a_n, \tag{15}$$

où Θ est une fonction fixe, arbitrairement prise, de ses arguments, qui satisfait à la condition

$$\frac{\partial^2 \Theta}{\partial \varphi \partial a_{n-1}} \not\equiv 0. \tag{16}$$

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LITTÉRATURE CITÉE

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MATHÉMATIQUES

SUR QUELQUES SOMMES DU TYPE DE VIGGO BRUN

Par W. TARTAKOWSKI

(Présenté par I. M. Vinogradow, de l'Académie, le 15. II. 1939)

Pour l'étude du crible approximatif d'Eratosthène on peut, comme l'a démontré V. Brun, utiliser les sommes du type:

$$1-\sum_{i_1\in\Omega_1}\frac{1}{p_{i_1}}+\sum_{(i_1,i_2)\in\Omega_2}\frac{1}{p_{i_1}p_{i_2}}-\ldots+(-1)^m\sum_{(i_1,\ldots,i_m)\in\Omega_m}\frac{1}{p_{i_1}p_{i_2}\cdots p_{i_m}} \qquad (1)$$

avec les domaines de sommation diverses $\Omega_1, \Omega_2, \ldots, \Omega_m$. Dans cette Note je vais examiner quelques sommes spéciales de cette espèce, formées de la manière suivante.

Soit $(0, z)$ un intervalle, où z est un nombre assez grand. Nous partageons cet intervalle en parties à l'aide des points:

$$t_0=z^{\alpha_0};\ t_1=z^{\alpha_0\alpha_1};\ t_2=z^{\alpha_0\alpha_1\alpha_2};\ \ldots\ t_n=z^{\alpha_0\alpha_1\ldots\alpha_n}=A,$$

où les nombres $\alpha_0, \alpha_1, \ldots \alpha_n$ sont des fractions positives, et le nombre A est contenu entre deux constantes fixées, mais suffisamment grandes. Tous les nombres premiers de l'intervalle $(0, z^{\alpha_0})$ sont énumerés en ordre d'accroissement (la quantité des nombres premiers qui ne dépassent pas A est désignée par g):

$$p_1=2;\ p_2=3;\ p_3=5;\ \ldots;\ p_g\leqslant A<p_{g+1}<p_{g+2}<\ldots<p_r\leqslant z^{\alpha_0}. \qquad (2)$$

Soit p_{r_ν} le nombre premier, le plus grand parmi ceux qui ne dépassent pas t_ν. Ainsi:

$$p_{r_0}=p_r\,(r_0=r) \quad \text{et} \quad p_{r_n}=p_g\,(r_n=g).$$

La partie de la série (2) contenue dans l'intervalle (A, z^{α_0}) est partagée par les nombres $p_{r_n}, p_{r_{n-1}}, \ldots, p_{r_1}, p_{r_0}$ en n segments. L'ensemble des éléments inverses d'un tel segment, c'est-à-dire: $\frac{1}{p_{r_\nu+1}}, \frac{1}{p_{r_\nu+2}}, \ldots, \frac{1}{p_{r_{\nu-1}}}$ sera désigné par $\bar{\sigma}_\nu$, et nous allons construire à l'aide de ces segments

$\overline{\sigma_\nu}(\nu=\lambda+1, \lambda+2, \ldots, \mu)$ un «schème en forme d'escalier du type de V. Brun» (1).

$$\left.\begin{matrix}\overline{\sigma}_\mu, \overline{\sigma}_{\mu-1}, \ldots, \overline{\sigma}_{\lambda+3}, \overline{\sigma}_{\lambda+2}, \overline{\sigma}_{\lambda+1} \\ \overline{\sigma}_\mu, \overline{\sigma}_{\mu-1}, \ldots, \overline{\sigma}_{\lambda+3}, \overline{\sigma}_{\lambda+2}, \overline{\sigma}_{\lambda+1}\end{matrix}\right\} k \text{ lignes}$$

$$\left.\begin{matrix}\overline{\sigma}_\mu, \overline{\sigma}_{\mu-1}, \ldots, \overline{\sigma}_{\lambda+3}, \overline{\sigma}_{\lambda+2} \\ \overline{\sigma}_\mu, \overline{\sigma}_{\mu-1}, \ldots, \overline{\sigma}_{\lambda+3}, \overline{\sigma}_{\lambda+2}\end{matrix}\right\} 2 \text{ lignes}$$

$$\left.\begin{matrix}\overline{\sigma}_\mu, \overline{\sigma}_{\mu-1}, \ldots, \overline{\sigma}_{\lambda+3} \\ \overline{\sigma}_\mu, \overline{\sigma}_{\mu-1}, \ldots, \overline{\sigma}_{\lambda+3}\end{matrix}\right\} 2 \text{ lignes} \qquad \text{(B)}$$

$$\ldots\ldots\ldots$$

$$\ldots\ldots\ldots$$

$$\left.\begin{matrix}\overline{\sigma}_\mu, \overline{\sigma}_{\mu-1} \\ \overline{\sigma}_\mu, \overline{\sigma}_{\mu-1}\end{matrix}\right\} 2 \text{ lignes}$$

$$\left.\begin{matrix}\overline{\sigma}_\mu \\ \overline{\sigma}_\mu\end{matrix}\right\} 2 \text{ lignes.}$$

Le schème est composé de $2(\mu-\lambda-1)+k$ lignes. Les premiers k lignes contiennent tous les nombres des segments $\overline{\sigma}_\mu, \ldots, \overline{\sigma}_{\lambda+1}$, les deux lignes suivantes se raccourcissent du côté droit d'un segment, les deux lignes suivantes se raccourcissent du côté droit d'un segment de plus, etc.; ainsi les deux dernières lignes contiennent chacune un segment unique. Les éléments du schème sont les nombres $\frac{1}{p}$.

Nous faisons correspondre à ce schème la somme suivante:

$$\sum_{\nu=0}^{2(\mu-\lambda-1)+k} \sum_{(i_1,\ldots,i_\nu)\in Q_\nu} \frac{(-1)^\nu}{p_{i_1} p_{i_2} \cdots p_{i_\nu}}.$$

Cette somme contient les produits (pris avec les signes correspondants) de toutes les combinaisons possibles d'éléments du schème (B), qui répondent aux deux conditions suivantes: 1) le μ-ième facteur de chaque produit $\left(\text{c'est-à-dire } \frac{1}{p_{i_\mu}}\right)$ est pris de la μ-ième ligne du schème (B), c'est-à-dire:

$$i_1 \leqslant r_0; i_2 \leqslant r_0; \ldots; i_k \leqslant r_0; i_{k+1} \leqslant r_1; i_{k+2} \leqslant r_1; i_{k+3} \leqslant r_2;$$
$$i_{k+4} \leqslant r_2; \ldots; i_\mu \leqslant r_{\left[\frac{\mu-k+1}{2}\right]} \quad (\mu = k, k+1, \ldots \nu);$$

2) les indices des facteurs de chaque produit forment une suite décroissante: $i_1 > i_2 > \ldots > i_\nu$.

Désignons la somme ainsi composée par le symbole:

$$P_k(\overline{\sigma}_{\lambda+1}, \overline{\sigma}_{\lambda+2}, \ldots, \overline{\sigma}_\mu).$$

Introduisons les notations:

$$\sigma_\nu = \sum_{r_\nu < i \leqslant r_{\nu-1}} \frac{1}{p_i}; \quad \sigma_\nu^{(s)} = \sum_{\substack{i_1>i_2>\ldots>i_s \\ r_\nu < i_k \leqslant r_{\nu-1}(k=1,2,\ldots s)}} \frac{1}{p_{i_1} p_{i_2} \cdots p_{i_s}},$$

$$\pi_\nu = \prod_{r_\nu < i \leqslant r_{\nu-1}} \left(1 - \frac{1}{p_i}\right).$$

Si $s > \rho_\nu = r_{\nu-1} - r_\nu$, alors $\sigma_\nu^{(s)} = 0$. Notons que $\sigma_\nu^{(1)} = \sigma_\nu$. D'après les formules connues de Mertens (2):

$$\sigma_\nu = \ln\frac{1}{\alpha_\nu} + \Theta\frac{5(1+\alpha_\nu)}{\ln z^{\alpha_0\alpha_1\ldots\alpha_\nu}} \quad \text{et} \quad \pi_\nu = \alpha_\nu e^{\Theta\frac{7(1+\alpha_\nu)}{\ln z^{\alpha_0\alpha_1,\ldots\alpha_\nu}}}.$$

Ici et dans la suite nous désignerons par Θ des nombres arbitraires, tels que $|\Theta| < 1$, et par $\overline{\Theta}$, tels que: $0 \leqslant \overline{\Theta} \leqslant 1$. Comme $5(1+\alpha_\nu) < 10$ et $7(1+\alpha_\nu) < 14$, on peut indiquer une constante F telle que:

$$\sigma_\nu = \ln\frac{1}{\alpha_\nu} + \frac{\Theta F}{\alpha_0\alpha_1\ldots\alpha_\nu \ln z};$$

$$\pi_\nu = \alpha_\nu\left[1 + \frac{\Theta F}{\alpha_0\alpha_1\ldots\alpha_\nu \ln z}\right].$$

Puisque $z^{\alpha_0\alpha_1\ldots\alpha_\nu} \geqslant A$ (car $\nu \leqslant n$), on a

$$\frac{\Theta F}{\alpha_0\alpha_1\ldots\alpha_\nu \ln z} \leqslant \frac{F}{\ln A} < \varepsilon,$$

où ε est une quantité positive arbitrairement petite, si A est suffisamment grand, et alors:

$$\sigma_\nu = \ln\frac{1}{\alpha_\nu} + \Theta\varepsilon < \ln\frac{1}{\alpha_\nu} + \varepsilon; \quad \pi_\nu = \alpha_\nu(1+\Theta\varepsilon) > \alpha_\nu(1-\varepsilon).$$

Nous allons choisir α_ν de telle façon que tous les σ_ν soient plus petits que 1 (c'est-à-dire $\alpha_\nu > e^{-(1-\varepsilon)}$).

Après ces remarques passons au calcul du symbole $P_k(\overline{\sigma}_{\lambda+1}, \ldots \overline{\sigma}_\mu)$. Ce symbole satisfait évidemment à la relation:

$$P_k(\overline{\sigma}_{\lambda+1}, \overline{\sigma}_{\lambda+2}, \ldots, \overline{\sigma}_\mu) = P_{k+2}(\overline{\sigma}_{\lambda+2}, \ldots \overline{\sigma}_\mu) - \overline{\sigma}_{\lambda+1}^{(1)} P_{k+1}(\overline{\sigma}_{\lambda+2}, \ldots, \overline{\sigma}_\mu) +$$
$$+ \overline{\sigma}_{\lambda+1}^{(2)} P_k(\overline{\sigma}_{\lambda+2}, \ldots, \overline{\sigma}_\mu) - \ldots + (-1)^k \overline{\sigma}_{\lambda+1}^{(k)} P_2(\overline{\sigma}_{\lambda+2}, \ldots, \overline{\sigma}_\mu).$$

Le calcul de $P_k(\overline{\sigma}_\mu)$ est très simple:

$$P_k(\overline{\sigma}_\mu) = 1 - \sigma_\mu^{(1)} + \sigma_\mu^{(2)} - \ldots + (-1)^k \sigma_\mu^{(k)} =$$
$$= \pi_\mu - (-1)^{k+1}[\sigma_\mu^{(k+1)} - \sigma_\mu^{(k+2)} + \ldots + (-1)^{\rho_\mu - k - 1}\sigma_\mu^{(\rho_\mu)}].$$

Mais $\sigma_\mu \sigma_\mu^{(s-1)} > s\sigma_\mu^{(s)}$, si $\sigma_\mu^{(s-1)} \neq 0$. D'où l'on tire trois inégalités pour $\sigma_\mu^{(s)}$:

$$\sigma_\mu^{(s)} < \frac{\sigma_\mu}{s}\sigma_\mu^{(s-1)}; \quad \sigma_\mu^{(s)} < \sigma_\mu^{(s-1)}; \quad \sigma_\mu^{(s)} < \frac{\sigma_\mu^s}{s!}.$$

Alors:

$$\sigma_\mu^{(k+1)} > \sigma_\mu^{(k+1)} - \sigma_\mu^{(k+2)} + \ldots + (-1)^{\rho_\mu - k - 1}\sigma_\mu^{(\rho_\mu)} > 0;$$

$$P_k(\sigma_\mu) = \pi_\mu + (-1)^k \overline{\Theta}\sigma_\mu^{k+1} = \pi_\mu + (-1)^k \overline{\Theta}\frac{\sigma_\mu^{k+1}}{(k+1)!}.$$

Ensuite:

$$P_k(\bar\sigma_{\mu-1}, \bar\sigma_\mu) = P_{k+2}(\bar\sigma_\mu) - \sigma_{\mu-1}^{(1)} P_{k+1}(\bar\sigma_\mu) + \ldots +$$

$$+ (-1)^k \sigma_{\mu-1}^{(k)} P_2(\bar\sigma_\mu) = \sum_{\nu=0}^{k} (-1)^\nu \sigma_{\mu-1}^{(\nu)} \left[\pi_\mu + \overline{\Theta}(-1)^{k-\nu} \frac{\sigma_\mu^{k+2-\nu+1}}{(k+2-\nu+1)!} \right] =$$

$$= \pi_\mu \sum_{\nu=0}^{k} (-1)^\nu \sigma_{\mu-1}^{(\nu)} + (-1)^k \overline{\Theta} \sum_{\nu=0}^{k} \frac{\sigma_{\mu-1}^{\nu} \sigma_\mu^{k-\nu+3}}{\nu!\,(k-\nu+3)!} =$$

$$= \pi_\mu \left[\pi_{\mu-1} + (-1)^k \overline{\Theta} \frac{\sigma_{\mu-1}^{k+1}}{(k+1)!} \right] + (-1)^k \overline{\Theta} \frac{\sigma_{\mu-1} + \sigma_\mu^{k+3}}{(k+3)!} =$$

$$= \pi_{\mu-1}\pi_\mu + (-1)^k \overline{\Theta} \left[\frac{\sigma_{\mu-1}^{k+1}}{(k+1)!} \pi_\mu + \frac{(\sigma_{\mu-1} + \sigma_\mu)^{k+3}}{(k+3)!} \right].$$

Continuant par la même méthode, nous obtenons:

$$P_k(\bar\sigma_1, \bar\sigma_2, \ldots, \bar\sigma_\mu) = \pi_1\pi_2 \ldots \pi_\mu + (-1)^k \overline{\Theta} \left[\frac{\sigma_1^{k+1}}{(k+1)!} \pi_2\pi_3 \ldots \pi_\mu + \right.$$

$$+ \frac{(\sigma_1+\sigma_2)^{k+3}}{(k+3)!} \pi_3 \ldots \pi_\mu + \ldots + \frac{(\sigma_1 + \ldots + \sigma_{\mu-1})^{k+2\mu-3}}{(k+2\mu-3)!} \pi_\mu +$$

$$\left. + \frac{(\sigma_1 + \ldots + \sigma_\mu)^{k+2\mu-1}}{(k+2\mu-1)!} \right],$$

où, si nous posons:

$$\frac{\sigma_1^{k+1}}{(k+1)!} \frac{1}{\pi_1} + \frac{(\sigma_1+\sigma_2)^{k+3}}{(k+3)!} \frac{1}{\pi_1\pi_2} + \ldots + \frac{(\sigma_1 + \ldots + \sigma_\mu)^{k+2\mu-1}}{(k+2\mu-1)!} \frac{1}{\pi_1\pi_2 \ldots \pi_\mu} = S_{k,\mu},$$

on a:

$$P_k(\bar\sigma_1, \bar\sigma_2, \ldots, \bar\sigma_\mu) = \pi_1, \pi_2 \ldots \pi_\mu [1 + (-1)^k \overline{\Theta} S_{k,\mu}].$$

Considérons encore le segment $\bar\sigma_{n+1}$, construit à l'aide de tous les nombres premiers de $p_1 = 2$ jusqu'à p_g; alors $\bar\sigma_{n+1}$ sera l'ensemble $\frac{1}{2}$, $\frac{1}{3}$, $\frac{1}{5}$, $\ldots$, $\frac{1}{p_g}$. Nous construisons le schème (B'), que nous pouvons obtenir du schème (B) avec $\lambda = 0$ et $\mu = n$, en y ajoutant du côté gauche encore une colonne de segments $\bar\sigma_{n+1}$. La différence entre le nombre de lignes dans cette colonne et dans la colonne précédente (la colonne des segments $\bar\sigma_n$) est égale à g.

La somme correspondant à ce schème sera désignée par $\overline{P}_k(\bar\sigma_1, \ldots, \bar\sigma_n; \bar\sigma_{n+1})$. Elle diffère de la somme $P_k(\bar\sigma_1, \ldots \bar\sigma_n)$ par le facteur:

$$\pi_{n+1} = \left(1 - \frac{1}{2}\right)\left(1 - \frac{1}{3}\right)\left(1 - \frac{1}{5}\right) \ldots \left(1 - \frac{1}{p_g}\right),$$

c'est-à-dire:

$$\overline{P}_k(\bar\sigma_1, \bar\sigma_2, \ldots, \bar\sigma_n; \bar\sigma_{n+1}) = \pi_1\pi_2 \ldots \pi_{n+1} \{1 + (-1)^k \overline{\Theta} S_{k,n}\}.$$

Désignons

$$\prod_{2 \leqslant p \leqslant k} \left(1 - \frac{1}{p}\right) = P(u)$$

(les p sont comme toujours des nombres premiers). Alors:

$$\overline{P}_k(\bar\sigma_1, \ldots, \bar\sigma_n; \bar\sigma_{n+1}) = P(z^{a_0}) \{1 + (-1)^k \overline{\Theta} S_{k,n}\}.$$

Quelquefois on a besoin de construire des sommes analogues en excluant de la série (2) quelques nombres premiers: $q_1, q_2, \ldots, q_\delta$. La série ainsi

obtenue sera désignée par (2′). Dans ce cas tout le calcul peut être reproduit avec la seule différence qu'on doit prendre à la place de $\pi_1, \pi_2, \ldots, \pi_{n+1}$ et $\sigma_1, \sigma_2, \ldots \sigma_{n+1}$ les quantités $\pi_1', \pi_2', \ldots \pi_{n+1}'$ et $\sigma_1', \sigma_2' \ldots \sigma_{n+1}'$, où

$$\sigma_\nu' = \sum_{\substack{r_\nu < i \leqslant r_{\nu-1} \\ p_i \neq q_j (j=1,2,\ldots,\delta)}} \frac{1}{p_i}; \quad \pi_\nu' = \prod_{\substack{r_\nu < i \leqslant r_{\nu-1} \\ p_i \neq q_j (j=1,2,\ldots,\delta)}} \left(1 - \frac{1}{p_i}\right)$$

On définit d'une manière semblable $\sigma_\nu'^{(s)}$, $\bar{\sigma}_\nu'$ et $\overline{P}_k(\bar{\sigma}_1', \bar{\sigma}_2', \ldots \bar{\sigma}_n'; \bar{\sigma}_{n+1}')$. Nous obtenons alors le résultat suivant:

$$\overline{P}_k(\bar{\sigma}_1', \bar{\sigma}_2', \ldots \bar{\sigma}_n'; \bar{\sigma}_{n+1}') = \pi_1' \pi_2' \ldots \pi_{n+1}' \left\{1 + (-1)^k \overline{\Theta} S_k'\right\}$$

où

$$S_k' = \frac{\sigma_1'^{k+1}}{(k+1)!} \frac{1}{\pi_1'} + \frac{(\sigma_1' + \sigma_2')^{k+2}}{(k+3)!} \frac{1}{\pi_1' \pi_2'} + \ldots + \frac{(\sigma_1' + \ldots + \sigma_n')^{k+2n-1}}{(k+2n-1)!} \frac{1}{\pi_1' \pi_2' \ldots \pi_n'};$$

$$\pi_1' \pi_2' \ldots \pi_{n+1}' = \frac{P(z^{\alpha_0})}{\left(1 - \frac{1}{q_1}\right)\left(1 - \frac{1}{q_2}\right) \ldots \left(1 - \frac{1}{q_\delta}\right)}.$$

Mais $\pi_i' \geqslant \pi_i$; $\frac{1}{\pi_i'} \leqslant \frac{1}{\pi_i}$; $\sigma_i' \leqslant \sigma_i$, d'où il suit que

$$S_k' \leqslant S_{k,n}.$$

Ainsi:

$$\overline{P}_k(\bar{\sigma}_1', \bar{\sigma}_2', \ldots, \bar{\sigma}_n'; \bar{\sigma}_{n+1}') = \frac{P(z^{\alpha_0})}{\left(1 - \frac{1}{q_1}\right) \ldots \left(1 - \frac{1}{q_\delta}\right)} \left\{1 + (-1)^k \overline{\Theta} S_k'\right\}.$$

Quand on donne à $\alpha_0, \alpha_1, \alpha_2, \ldots$ des valeurs numériques concrètes, il est facile d'évaluer une borne supérieure pour S_k'. Prenons par exemple: $\alpha_1 = \alpha_2 = \ldots = \alpha_n = \alpha$. Dans ce cas nous introduisons le symbole $B_k[z, \alpha_0, \alpha]$ à la place de $\overline{P}_k(\bar{\sigma}_1', \bar{\sigma}_2', \ldots, \bar{\sigma}_n'; \bar{\sigma}_{n+1}')$. Quand $\alpha = \frac{2}{3}$, nous obtenons:

$$S_k' \leqslant S_{k,n} \leqslant \sum_{s=1}^{n} \frac{\left[\varepsilon\left(\ln \frac{3}{2} + \varepsilon\right)\right]^{2s+k-1}}{(2s+k-1)!} \left(\frac{3}{2} + \varepsilon\right)^s.$$

Un calcul facile montre que dans ce cas:

$$S_1' < 0.193; \quad S_2' < 0.029; \quad S_3' < 0.0034 \text{ etc.}$$

De la formule de Mertens

$$P(z^{\alpha_0}) = \frac{e^{-c}}{\ln z^{\alpha_0}} \left(1 + \frac{7\Theta}{\ln z^{\alpha_0}}\right)$$

(où c est la constante eulerienne), on obtient la formule:

$$B_k[z, \alpha_0, \alpha] = \frac{e^{-c}}{\left(1 - \frac{1}{q_1}\right) \ldots \left(1 - \frac{1}{q_\delta}\right)} \cdot \frac{1}{\alpha_0 \ln z} \left\{1 + (-1)^k \overline{\Theta} S_k''\right\},$$

où la différence $S_k'' - S_k'$ est plus petite qu'un nombre positif arbitrairement petit, si z est suffisamment grand.

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MATHÉMATIQUES

LA MÉTHODE DU CRIBLE APPROXIMATIF «ÉLECTIF»

Par W.TARTAKOWSKI

(Présenté par I. M. Vinogradow, de l'Académie, le 15.II.1939)

Soit $N(\Delta, D; z; p_1, \ldots, p_r)$ le nombre des entiers positifs de la progression $Du+\Delta$ [où $(\Delta, D)=1$ et $u=0, \pm 1, \pm 2, \ldots$] qui ne surpassent pas z et n'appartiennent à aucune des progressions $p_1u_1, p_2u_2, \ldots, p_ru_r$. Ici $u_i = \pm 1, 2, 3, \ldots$; quant aux nombres premiers $p_1, p_2, \ldots, p_r$ ils satisfont à la condition: $(p_i, D)=1$ $(i=1, 2, \ldots, r)$ (1).

La relation suivante est évidente:

$$N(\Delta, D; z; p_1, \ldots, p_r) = N(\Delta, D; z; p_1, \ldots, p_{r-1}) - N(\Delta_r, Dp_r; z; p_1, \ldots p_{r-1}).$$

L'application réitérée de cette relation au second membre donne une relation nouvelle:

$$N(\Delta, D; z; p_1, \ldots, p_r) = N(\Delta, D; z) - \sum_{1 \leqslant i_1 \leqslant r} N(\Delta_{i_1}, Dp_{i_1}; z) + \sum_{1 \leqslant i_2 < i_1 \leqslant r} N(\Delta_{i_1 i_2}, Dp_{i_1}p_{i_2}; z; p_1 \ldots, p_{i_2-1}). \tag{1}$$

Si nous supprimons quelques termes dans la dernière somme du second membre, ce second membre n'augmentera pas, parce que les termes supprimés ne sont pas négatifs, et ainsi (1) deviendra une inégalité, qui permet d'évaluer $N(\Delta, D; z; p_1, \ldots, p_r)$ inférieurement. Nous désignons le domaine de sommation diminué par $\bar{\omega}_2$, et le domaine de sommation de i_1 $(1 \leqslant i_1 \leqslant r)$ sera désigné par $\bar{\omega}_1$. Alors

$$N(\Delta, D; z; p_1, \ldots, p_r) \geqslant N(\Delta, D; z) - \sum_{i_1 \in \bar{\omega}_1} N(\Delta_{i_1}, Dp_{i_1}; z) + \sum_{(i_1, i_2) \in \bar{\omega}_2} N(\Delta_{i_1 i_2}, Dp_{i_1}p_{i_2}; z; p_1, \ldots, p_{i_2-1}). \tag{2}$$

Nous allons prendre pour $\bar{\omega}_2$ la partie du domaine triangulaire: $1 \leqslant i_2 < i_1 \leqslant r$ borné par la condition supplémentaire: $i_2 < \omega_2(i_1)$. Appliquons aux membres de la dernière somme de (2) la même formule (2) et choisissons comme restriction pour i_4 la condition supplémentaire: $i_4 < \omega_4(i_1, i_2, i_3)$.

En réitérant ce procédé et en supprimant dans la dernière somme les membres non-négatifs, nous obtiendrons l'inégalité:

$$N(\Delta, D; z; p_1, \dots p_r) \geqslant \sum_{k=0}^{\eta} \Bigg[\sum_{\substack{(i_1, \dots i_{2k-1}) \in \overline{\omega}_{2k-1} \\ i_{2k} < \min[i_{2k-1}, \omega_{2k}(i_1, \dots, i_{2k-1})]}} N(\Delta_{i_1, \dots i_{2k}}, Dp_{i_1} \dots p_{i_{2k}}; z) - \sum_{\substack{(i_1, \dots, i_{2k}) \in \overline{\omega}_{2k} \\ i_{2k+1} < i_{2k}}} N(\Delta_{i_1 \dots i_{2k+1}}, Dp_{i_1} \dots p_{i_{2k+1}}; z) \Bigg]. \tag{3}$$

Ici $\omega_0 = r$; $\overline{\omega}_\nu$ est le domaine de sommation des symboles $N(\Delta_{i_1 \dots i_\nu}, Dp_{i_1} \dots p_{i_\nu}; z)$. Notons que:

$$N(\Delta', D'; z) = \frac{z}{D'} + \Theta', \text{ où } |\Theta'| < 1.$$

Parce que la somme algébrique des quantités Θ' pour tous les termes du second membre de (3) est inférieure au nombre de ces termes que nous désignerons par R, l'inégalité suivante dérive de la formule (3):

$$N(\Delta, D; z; p_1, \dots, p_r) > \frac{z}{D} \sum_{k=0}^{q} \Bigg[\sum_{\substack{(i_1, \dots, i_{2k-1}) \in \overline{\omega}_{2k-1} \\ i_{2k} < \min[i_{2k-1}, \omega_{2k}(i_1, \dots, i_{2k-1})]}} \frac{1}{p_{i_1} \dots p_{i_{2k}}} - \sum_{\substack{(i_1, \dots, i_{2k}) \in \overline{\omega}_{2k} \\ i_{2k+1} < i_{2k}}} \frac{1}{p_{i_1} \dots p_{i_{2k+1}}} \Bigg] - R. \tag{4}$$

Les fonctions $\omega_{2k}(i_1, \dots, i_{2k-1})$ peuvent être choisies de manières diverses.

Viggo Brun prenait pour la suite $p_1, p_2, \dots p_r$ * tous les nombres premiers de 2 jusqu'au plus grand nombre qui ne surpasse pas z^{α_0} (1) excepté les diviseurs premiers de D, que nous allons désigner par $q_1, \dots, q_\delta$. Il choisissait $\omega_{2k}(i_1, \dots, i_{2k-1}) = z^{\alpha_0 \alpha^k}$ $(k = 1, 2, \dots, n-1)$, où α_0 et α sont des fractions positives regulières et le nombre n est déterminé par la condition:

$$z^{\alpha_0 \alpha^{n-1}} \geqslant A_0 > z^{\alpha_0 \alpha^n},$$

où A_0 est assez grand, mais indépendant de z. Autrement dit:

$$n = \left[\frac{1}{\ln \frac{1}{\alpha}} \left(\ln \ln z - \ln \ln A + \ln \frac{1}{\alpha_0} \right) \right] + 1. \tag{5}$$

Le nombre $q = n + \left[\frac{g+1}{2} \right]$, où g est le nombre des nombres premiers ne surpassant pas A. Si $k \geqslant n$, $\omega_{2k} = z^{\alpha_0 \alpha^n}$. En choisissant les fonctions ω_k de cette manière, on trouve pour R l'évaluation suivante:

$$R \leqslant r\omega_2^2 \omega_4^2 \dots \omega_{2n}^2 2^g \leqslant \frac{z^{\alpha_0 (1 + 2\alpha + 2\alpha^2 + \dots) 2^g}}{\alpha_0 \ln z \, (\alpha_0 \alpha \ln z)^2 (\alpha_0 \alpha^2 \ln z)^2 \dots} \leqslant \frac{z^{\alpha_0 \frac{1+\alpha}{1-\alpha}}}{(\ln z)^\mu},$$

où μ est un nombre fixe positif arbitrairement grand, quand z est suffisamment grand. Si $\alpha = \frac{2}{3}$ et $\alpha_0 = \frac{1}{5}$, la formule (4) prend la forme:

$$N(\Delta, D; z; p_1, \dots, p_r) \geqslant \frac{z}{D} B_1 \left(z; \frac{1}{5}; \frac{2}{3} \right) -$$

$$- \frac{z}{(\ln z)^\mu} \geqslant \frac{5e^{-c}}{\varphi(D)} 0.807 \frac{z}{\ln z} - \frac{z}{(\ln z)^2}.$$

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* Nous allons dans la suite profiter des notations et des résultats de ma Note «Sur quelques sommes du type de Viggo Brun», v. cette issue.

Contrairement à la méthode exposée précédemment, nous considérons ω_{2k} comme une fonction de i_1. Ainsi pour chaque p_{i_1} le procédé suivant du crible approximatif sera individualisé. Un tel crible approximatif peut être nommé «électif au premier pas». La manière dont ω_{2k} dépend de i_1 sera différente pour $p_{i_1} \leqslant z^{\frac{1}{7}}$ et pout $p_{i_1} > z^{\frac{1}{7}}$. La suite $p_1, p_2, \ldots, p_r$ sera choisie comme précédemment, seulement α_0 sera posé égal à $\frac{1}{3}+\varepsilon$.

$$\text{Si } p_{i_1} \leqslant z^{\frac{1}{7}};\ \omega_{2k}(i_1) = z^{\frac{1}{7}}\left(\tfrac{2}{3}\right)^{k-1} \quad (k=1, 2, \ldots, n');$$

$$\omega_{2k}(i_1) = z^{\frac{1}{7}}\left(\tfrac{2}{3}\right)^{n'}, \quad (k=n'+1, \ldots).$$

$$\text{Si } z^{\frac{1}{7}} < p_{i_1} \leqslant z^{\frac{1}{3}+\varepsilon};\ \omega_{2k}(i_1) = \left(\frac{z}{p_{i_1}}\right)^{\frac{1}{6}}\left(\tfrac{2}{3}\right)^{k-1} \quad (k=1, 2, \ldots n'');$$

$$\omega_{2k}(i_1) = \left(\frac{z}{p_{i_1}}\right)^{\frac{1}{6}}\left(\tfrac{2}{3}\right)^{n''}, \quad (k=n''+1, \ldots).$$

Ici d'après (5)

$$n' = \left[\frac{1}{\ln\frac{3}{2}}(\ln\ln z - \ln\ln A + \ln 7)\right] + 1;$$

$$n'' = \left[\frac{1}{\ln\frac{3}{2}}\left(\ln\ln\frac{z}{p_{i_1}} - \ln\ln A + \ln 5\right) + 1\right].$$

Conformément à ce partage, la somme dans la partie droite de l'inégalité (4) sera divisée en deux sommes et le membre restant en deux membres restants:

$$Q_1 - Q_2 \text{ et } R_1 + R_2,$$

où

$$Q_1 = \frac{z}{D} B_3\left[z; \tfrac{1}{7}; \tfrac{2}{3}\right];\ Q_2 = \sum_{p>z^{\frac{1}{7}}}^{\leqslant z^{\frac{1}{3}+\varepsilon}} \frac{1}{p} B_2\left[\frac{z}{p}; \tfrac{1}{6}; \tfrac{2}{3}\right];\ R_1 \leqslant \frac{z}{(\ln z)^2};$$

$$R_2 \leqslant \sum_{p>z^{\frac{1}{7}}}^{\leqslant z^{\frac{1}{3}+\varepsilon}} \frac{\left(\frac{z}{p}\right)}{\left(\ln\frac{z}{p}\right)^2} \leqslant \frac{z}{\left(\ln\frac{z}{z^{\frac{1}{3}+\varepsilon}}\right)^2} \sum_{p>z^{\frac{1}{7}}}^{\leqslant z^{\frac{1}{3}+\varepsilon}} \frac{1}{p} \leqslant$$

$$\leqslant \frac{z}{\left(\frac{2}{3}-\varepsilon\right)^2 \ln^2 z} 2\ln\left(\frac{7}{3}+7\varepsilon\right) \leqslant 4\frac{z}{(\ln z)^2}.$$

Mais d'après ma Note déjà citée on a:

$$Q_2 = \sum_{p>z^{\frac{1}{7}}}^{\leqslant z^{\frac{1}{3}+\varepsilon}} \frac{z}{D}\frac{1}{p} \frac{e^{-c}}{\left(1-\frac{1}{q_1}\right)\ldots\left(1-\frac{1}{q_s}\right)} \frac{1}{\frac{1}{6}\ln\frac{z}{p}} \times$$

$$\times(1+\overline{\Theta}\cdot 0.029) \leqslant 1.029\frac{e^{-c}}{\varphi(D)}\frac{z}{\left(\frac{1}{6}\right)} \sum_{p>z^{\frac{1}{7}}}^{\leqslant z^{\frac{1}{3}+\varepsilon}} \frac{1}{p\ln\frac{z}{p}}.$$

Les transformations habituelles d'Abel (directes et inverses) à l'aide de la formule connue: $\sum_{p\leqslant x}\frac{1}{p}=\ln\ln x+B+O\left(\frac{1}{\ln^{\mu}x}\right)$(2) ($\mu$ a le sens défini plus haut) permettent de remplacer la dernière somme par une intégrale:

$$\sum_{p>z^{\frac{1}{7}}}^{\leqslant z^{\frac{1}{3}+\varepsilon}}\frac{1}{p\ln\frac{z}{p}}=\frac{1}{\ln z}\int_{\frac{1}{7}}^{\frac{1}{3}+\varepsilon}\frac{du}{u(1-u)}+O\left(\frac{1}{\ln^2 z}\right)=\frac{1}{\ln z}\ln(3+\varepsilon')+O\left(\frac{1}{\ln^2 z}\right).$$

Comme $D\left(1-\frac{1}{q_1}\right)\ldots\left(1-\frac{1}{q_\delta}\right)=\varphi(D)$, nous obtiendrons:

$$N(\Delta, D; z; p_1, \ldots p_r)\geqslant e^{-c}(1-0.0034)\frac{z}{\varphi(D)}\times$$

$$\times\frac{1}{\frac{1}{7}\ln z}-e^{-c}\cdot 1.029\frac{z}{\varphi(D)}\frac{1}{\frac{1}{6}\ln z}+O\left(\frac{z}{\ln^2 z}\right)\geqslant 0.45\frac{z}{\varphi(D)\ln z}+O\left(\frac{z}{\ln^2 z}\right).$$

Il en suit que dans chaque progression $Du+\Delta$ [où $(D, \Delta)=1$], si z est suffisamment grand, il existe au moins $a\frac{z}{\varphi(D)\ln z}$ (a est une constante absolue) nombres premiers inférieurs à z ou bien de nombres inférieurs à z et ayant deux facteurs premiers, qui doivent être renfermés entre $z^{\frac{1}{3}+\varepsilon}$ et $z^{\frac{2}{3}-2\varepsilon}$. L'intervalle peut être fortement serré, si l'on se sert de ma méthode d'élection aux troisième, cinquième etc. pas du crible approximatif. J'ai calculé que par cette méthode le résultat démontré plus haut peut être précisé comme il suit:

Théorème. *Dans la progression* $Du+\Delta$ [*où* $(D, \Delta)=1$] *dans l'intervalle* $(0, z)$, *si* z *est suffisamment grand* [$z\geqslant z_0(D)$], *il existe au moins* $a\frac{z}{\varphi(D)\ln z}$ (*a est une constante absolue qui ne dépend ni de* D, *ni de* Δ, *ni de* z) *nombres premiers ou bien de nombres ayant deux facteurs premiers, qui doivent être renfermés entre* $z^{\frac{1}{2}-\delta}$ *et* $z^{\frac{1}{2}+\delta}$, *où* $\delta=0.01$.

Le nombre δ peut être encore diminué.

Les premières applications de la méthode du crible approximatif électif à l'équation d'Euler et aux problèmes des nombres jumeaux montrent qu'il existe des solutions de l'équation d'Euler en nombres «presque premiers» et qu'il existe une infinité de paires de nombres jumeaux «presque premiers» et d'ailleurs ces nombres presque premiers ont au plus quatre facteurs premiers.

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THEORY OF ELASTICITY

ON THE DURABILITY OF THICK-WALLED CYLINDERS

By I. M. GORDON

(Communicated by B. G. Galerkin, Member of the Academy, 25. II. 1939)

In connexion with the papers of Beliaev and Sinitzki (1) on the question of theoretical investigation of stresses and deformations in thick-walled cylinders, it may be useful to communicate some data obtained by the author at the investigation of thick-walled cylinders under the conditions of super-high pressures, without touching the methodical side of the investigations which will be discussed elsewhere.

As is known, the theory of autofastening (autofrettage) assumes that the fundamental factors in the effection of autofastening are: a) the rise of the limit of elasticity due to the phenomenon of cold riveting, b) the creation in the walls of the cylinders of preliminary useful stresses of tangential contraction. The influence of the first factor is confined to comparatively narrow limits (in comparison with the pressures which it is possible to obtain in cylinders), since the rise of the original limit of elasticity can be thought to be possible only up to the value of temporary resistance.

The existing practice of the autofrettage process deals with cylinders, the ratio of the diameters of which does not exceed 3 ($\frac{D}{d} \leqslant 3$) and with pressures which just exceed the numerical value of the limit of fluidity and are aimed at the rise of the limit of elasticity.

The question on the residual stresses considered within these limits ($\frac{D}{d} \leqslant 3$ and pressures at the autofrettage process) is up to a certain extent solved in the papers of Macrae, Malowal, Drosdov and in the last paper of Beliaev and Sinitzki. However, the question on the durability of thick-walled cylinders with the ratio of diameters $\frac{D}{d} > 5$ under the conditions of super-high pressures, exceeding several times the numerical value of temporary resistance, has not been yet investigated to its full extent.

Therefore, the problem which has been considered by the author consisted in widening of the limits of experimental investigations of thick-walled cylinders with the ratio of diameters $\frac{D}{d} > 5$ and pressures, considerably exceeding the numerical value of temporary resistance, but not leading to the destruction of the cylinder, and explained by this manner under these circumstances the significance of the residual stresses in the general process of fastening of cylinders. The diapason of pressures of investigated cylinders was from 8 500 to 22 000 atm.

Thus, in the present paper is given a short account of an investigation which we may formulate as the investigation of residual stresses in steel thick-walled cylinders made under super-high pressure.

The object of investigation: thick-walled cylinders with the outer diameter 100 mm and the inner diameter 15 mm. Material: carbon steel 0.45% C, the limit of fluidity 34 kg/mm² and temporary resistance 65—72 kg/mm².

For the present investigation a method of obtaining super-high pressure was especially developed, which enabled us to obtain pressures in bottomless cylinders. The cylinders were made under pressures 8 500, 9 100, 11 000, 16 000, 19 300, 20 000 and 22 000 atm. The thus obtained cylinders were then subjected to a further investigation, consisting in determination in them of residual stresses in the tangential and the radial directions. From the existing methods of determination of residual stresses the most adequate in our conditions proved to be the method of Sachs.

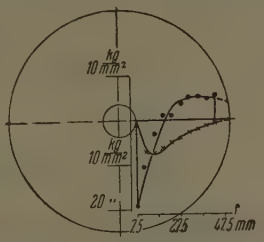

Fig. 1. Cylinder treated under pressure of 8 500 atm.

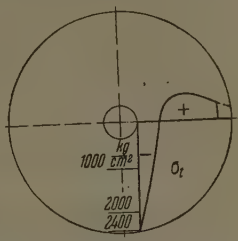

Fig. 2. Cylinder treated under pressure of 9 100 atm.

In order to establish the accuracy of the obtained results concerning the residual stresses, a certain control was carried out for each cylinder separately. This control consisted in confirming the existence of the following conditions of equilibration $\Sigma \sigma_t \Delta F = 0$ and $\Sigma \sigma_t \Delta F . x = 0$, that is the resultant and moment of the residual stresses must be equal to zero. In all cases these conditions were satisfied.

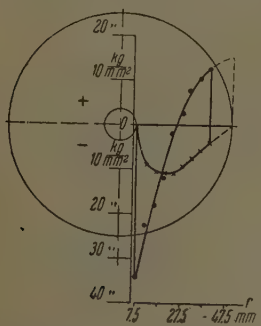

Fig. 3. Cylinder treated under pressure of 16 000 atm.

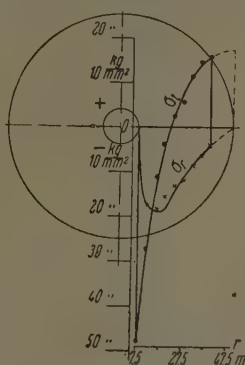

Fig. 4. Cylinder treated under pressure of 20 000 atm.

The results of the investigations are given in the following Figs. In Figs. 1, 2, 3 and 4 are given the graphs of the distribution of the residual stresses σ_t and σ_r in cylinders made under the pressures of 8 500, 9 100, 16 100 and 20 500 atm. respectively. From these graphs it follows that the maximum residual stresses σ_t increase with the pressure up to 20 500 atm.

The repeated tests of cylinders made under pressures of one and the same order show a satisfactory coincidence of residual stresses up to 20 500 atm. The character of the change of residual stresses showing the order of their increase with the rise of the pressure remains one and the same in all cases and shows their importance in the conditions of super-high pressures.

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MATHEMATISCHE PHYSIK

DIE GREENSCHE FUNKTION DER WÄRMELEITUNGSGLEICHUNG

Von P. V. SOLOVIEFF

(Vorgelegt von S. L. Sobolev, Mitglied der Akademie, d.28.II.1939)

In der Ebene xoy sei ein Gebiet D gegeben, das von zwei der Achse ox parallelen Geraden ($y=0$, $y=h$) und den Geraden

$$(S_1) \quad x=k_1\eta$$

$$(S_2) \quad x=k_2\eta+b \qquad \left(k_1,\ k_2 \neq \pm\infty,\ h<\frac{b}{2(k_1-k_2)}>0\right)$$

beschränkt ist. Ein solches Gebiet nennen wir ein Trapez. Wir konstruieren die Greensche Funktion (1) der Gleichung

$$\frac{\partial^2 u}{\partial x^2}-\frac{\partial u}{\partial y}=0 \tag{1}$$

für ein Trapez. Die Greensche Funktion werden wir in der Form der Reihe

$$G(x, y; \xi, \eta)=\sum_{n=1}^{\infty}\Psi_n(x, y; \xi, \eta) \tag{2}$$

suchen, wo die Funktion $\Psi_n(x, y; \xi, \eta)$ $(n=1, 2, \ldots)$ in Bezug auf die Veränderlichen (x, y) eine Lösung der Gleichung (1) und in Bezug auf die Veränderlichen (ξ, η) eine Lösung der Gleichung

$$\frac{\partial^2 v}{\partial \xi^2}-\frac{\partial v}{\partial \eta}=0 \tag{3}$$

ist.

Ferner, falls n eine gerade Zahl ist, soll

$$\Psi_{n+1}(x, y; k_1\eta, \eta)=-\Psi_n(x, y; k_1\eta, \eta)$$

sein, und falls n eine ungerade Zahl ist,

$$\Psi_{n+3}(x, y; k_2\eta+b, \eta)=-\Psi_n(x, y; k_2\eta+b, \eta).$$

Die Funktion $\Psi_1(x, y; \xi, \eta)$ auf (S_1) und die Funktion $\Psi_2(x, y; \xi, \eta)$ auf S_2 sollen Werte annehmen, die mit den Werten der Funktion

$$\frac{1}{\sqrt{y-\eta}}e^{-\frac{(x-\xi)^2}{4(y-\eta)}} \tag{3'}$$

auf diesen Abschnitten übereinstimmen. Man überzeugt sich leicht, dass solche Funktionen $\Psi_n(x, y; \xi, \eta)$ durch

$$\Psi_1(x, y; \xi, \eta) = \frac{1}{\sqrt{y-\eta}} e^{-\frac{(2k_1 y - x - \xi)^2}{4(y-\eta)} - k_1(x - k_1 y)}$$

$$\Psi_2(x, y; \xi, \eta) = \frac{1}{\sqrt{y-\eta}} e^{-\frac{(2k_2 y + 2b - x - \xi)^2}{4(y-\eta)} - \{k_2 x - k_2^2 y - k_2 b\}}$$

$$\Psi_3(x, y; \xi, \eta) = -\frac{1}{\sqrt{y-\eta}} e^{-\frac{[2k_1 y - 2(k_2 y + b) + x - \xi]^2}{4(y-\eta)} - \{(k_1 - k_2)x - (k_1 - k_2)^2 y + (2k_1 - k_2)b\}}$$

u. s. w. gegeben sind.

Wenn wir diese Funktionen in die Reihe (2) einsetzen, so erhalten wir

$$G(x, y; \xi, \eta) =$$

$$= \frac{1}{\sqrt{y-\eta}} \sum_{n=1}^{\infty} \left\{ e^{-\frac{[2nk_1 y - 2(n-1)(k_2 y + b) - x - \xi]^2}{4(y-\eta)} - \{[nk_1 - (n-1)k_2]x - [nk_1 - (n-1)k_2]^2 y + (n-1)[nk_1 - (n-1)k_2]b\}} + \right.$$

$$+ e^{-\frac{[2n(k_2 y + b) - 2(n-1)k_1 y - x - \xi]^2}{4(y-\eta)} - \{[nk_2 - (n-1)k_1]x - [nk_2 - (n-1)k_1]^2 y - n[nk_2 - (n-1)k_1]b\}} -$$

$$- e^{-\frac{[2nk_1 y - 2n(k_2 y + b) + x - \xi]^2}{4(y-\eta)} - \{n(k_2 - k_1)x - n^2(k_2 - k_1)^2 y + n[(n+1)k_1 - nk_2]b\}} -$$

$$\left. - e^{-\frac{[2nk_1 y + 2n(k_2 y + b) + x - \xi]^2}{4(y-\eta)} - \{n(k_1 - k_2)x - n^2(k_1 - k_2)^2 y + n[(n+1)k_1 - nk_2]b\}} \right\}. \quad (4)$$

Die Reihe (4) konvergiert im Gebiete D für $0 \leqslant y \leqslant h < \frac{b}{2(k_1 - k_2)}$ absolut und gleichmässig, da sie als Majorante die konvergente Reihe

$$\sum_{n=1}^{n_0-1} \Psi_n(x, y; \xi, \eta) + \frac{4}{\sqrt{y-\eta}} \sum_{n=n_0}^{\infty} e^{-kn^2} \qquad (k > 0)$$

hat.

Hier sind die Zahlen n_0 und k so gewählt, dass für $n \geqslant n_0$

$$B_n = \min \left\{ \left\{ \frac{\left[2k_1 y - 2\left(1 - \frac{1}{n}\right)(k_2 y + b) - \frac{x + \xi}{n}\right]^2}{4(y-\eta)} + \left[\frac{nk_1 - (n-1)k_2}{n^2} x - \right.\right.\right.$$

$$\left. - \left(k_1 - \left(1 - \frac{1}{n}\right)k_2\right)^2 y + \left(1 - \frac{1}{n}\right)\left(k_1 - \left(1 - \frac{1}{n}\right)k_2\right) b\right] \Bigg\};$$

$$\left\{ \frac{\left[2(k_2 y + b) - 2\left(1 - \frac{1}{n}\right)k_1 y + \frac{x + \xi}{n}\right]^2}{4(y-\eta)} + \left[\frac{nk_2 - (n-1)k_1}{n^2} x - \right.\right.$$

$$\left. - \left(k_2 - \left(1 - \frac{1}{n}\right)k_1\right)^2 y - \left(k_2 - \left(1 - \frac{1}{n}\right)k_1\right) b\right] \Bigg\};$$

$$\left\{ \frac{\left[2k_1 y - 2(k_2 y + b) + \frac{x - \xi}{n}\right]^2}{4(y-\eta)} + \left[\frac{k_2 - k_1}{n} x - (k_2 - k_1)^2 y + \right.\right.$$

$$\left. + \left(\left(1 + \frac{1}{n}\right)k_1 - k_2\right) b\right] \Bigg\};$$

$$\left\{\frac{\left[-2k_1y+2(k_2y+b)-\frac{x-\xi}{n}\right]^2}{4(y-\eta)}+\left[\frac{k_1-k_2}{n}x-(k_1-k_2)^2y+\right.\right.$$

$$\left.\left.+\left(\left(1+\frac{1}{n}\right)k_1-k_2\right)b\right]\right\}\geqslant k>0$$

ausfällt, was möglich ist, weil

$$\lim_{n\to\infty} B_n>0$$

ist.

Ebenso leicht ist es die absolute und gleichmässige Konvergenz im Gebiete D der Reihen zu beweisen, die wir mit

$$\frac{\partial g}{\partial x};\ \frac{\partial g}{\partial \xi};\ \frac{\partial^2 g}{\partial \xi^2};\ \frac{\partial^2 g}{\partial x^2};\ \frac{\partial g}{\partial y};\ \frac{\partial g}{\partial \eta};$$

bezeichnen und die aus der Reihe (4) durch gliedweise Differentiation erhalten werden. Daraus folgt

$$\left(\frac{\partial^2}{\partial x^2}-\frac{\partial}{\partial y}\right)G(x,y;\xi,\eta)=\sum_{n=1}^{\infty}\left\{\left(\frac{\partial^2}{\partial x^2}-\frac{\partial}{\partial y}\right)\Psi_n(x,y;\xi,\eta)\equiv 0\right\}\equiv 0$$

und

$$\left(\frac{\partial^2}{\partial \xi^2}+\frac{\partial}{\partial \eta}\right)G(x,y;\xi,\eta)=\sum_{n=1}^{\infty}\left\{\left(\frac{\partial^2}{\partial \xi^2}+\frac{\partial}{\partial \eta}\right)\Psi_n(x,y;\xi,\eta)\equiv 0\right\}\equiv 0.$$

Ferner sind die Werte der Funktion $G(x,y;\xi,\eta)$ auf (S_1) gleich

$$G(x,y;k_1\eta,\eta)=\frac{1}{\sqrt{y-\eta}}e^{-\frac{(x-k_1\eta)^2}{4(y-\eta)}}$$

weil nach der Substitution $\xi=k_1\eta$ die Exponenten der Potenzen der zweiten und des dritten Summanden des allgemeinen Gliedes der Reihe (4) und die der Potenzen des vierten und des ersten Summanden übereinstimmen, wenn mann n durch $n+1$ ersetzt. Ebenso durch die Substitution $\xi=k_2\eta+b$ überzeugen wir uns, dass die Werte der Funktion $G(x,y;\xi,\eta)$ auf (S_2) mit den Werten der Funktion (3′) übereinstimmen.

Aus dem Gesagten folgt, dass die konstruierte Funktion $G(x,y;\xi,\eta)$ die Greensche Funktion der Gleichung (1) für das Trapez ist.

Setzt man in der Formel (4) $k_1=k_2=0$, so bekommt man die Greensche Funktion für das Rechteck.

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PHYSIQUE

L'OBSERVATION DES NOYAUX DE RECUL AU BOMBARDEMENT DE L'URANIUM PAR LES NEUTRONS

Par A. JDANOFF et L. MYSSOWSKY

(Présenté par W. G. Khlopine, de l'Académie, 7. III. 1939)

Il a été établi par les travaux de Hahn et Strassmann (1), Joliot (2), Fowler et Dodson (3), ainsi que par ceux de Frisch (4), qu'au bombardement de l'uranium et du thorium par les neutrons des noyaux de recul se produisent dont le numéro atomique est d'environ 50 et l'énergie (d'après les données de Frisch) atteint 50 meV. Joliot (5) indique dans son article que des noyaux de recul de très grande énergie ont été observés auparavant dans la chambre de Wilson à faible pression, mais alors il fut encore impossible de donner une juste explication à ce phénomène.

×600.

Nous avons tenté d'obtenir des traces de noyaux de recul sur des plaques photographiques à une couche épaisse d'émulsion. Les expériences ont été effectuées d'une manière suivante: une couche d'uranate de sodium a été mise dans une cavité quadrangulaire, faite dans une brique de paraffine d'épaisseur d'un cm. Au-dessus de la couche d'uranate était placée une plaque photographique recouverte d'une couche épaisse d'émulsion, tournée vers l'uranate. Une source de neutrons de 280 mC se trouvait sous la paraffine à une distance de 5 cm de la couche d'uranate. Entre la source des neutrons et la paraffine un écran d'épaisseur de 3 cm était placé. Cet écran servait à diminuer le fond causé par les rayons γ. L'exposition de 40 min. a été trouvée la plus convenable. A la même exposition furent soumises les plaques de contrôle qui furent exposées une fois à l'uranate de sodium, mais sans source de neutrons, et une autre fois seulement à la source de neutrons, sans l'uranate. Toutes les autres conditions de l'expérience n'étaient pas changées pendant l'expérience de contrôle.

A l'examen stéréoscopique sous microscope rien ne fut trouvé sur les plaques de contrôle, outre le fond des rayons γ, des particules α et des protons de recul. Quand au plaques soumises à l'action des neutrons en présence d'uranate de sodium, des traces furent trouvées qui en épaisseur et en densité de la disposition des grains dépassaient les traces des particules α au moins deux ou trois fois.

La figure représente la photographie stéréoscopique des deux particules qui, par hasard, se trouvent situées l'une près de l'autre. Là se trouve

aussi une trace des particules α auxquelles on peut les comparer. La présence de ces traces, si épaisses en comparaison avec les traces des particules α, peut être attribuée naturellement aux noyaux de recul, surgissant à la désintégration de l'uranium. En comptant toutes les particules, les courtes ainsi que les longues aux parcours de 0.5 à 2.0 cm dans l'air, on peut voir que toute la plaque dont la surface est de 6 cm^2 recevra au cours de 40 min. à peu près 6000 de ces particules, ce qui fait à peu près 150 particules par minute. Si l'on ne prend en considération que les longs parcours, ce nombre sera réduit plusieurs fois et se rapprochera au nombre trouvé par Frisch (15 par minute). Ces expériences seront répétées prochainement avec une émulsion spécialement choisie pour de différents éléments lourds.

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Académie des Sciences de l'URSS.
Leningrad.

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le 10. III. 1939.

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PHYSICS

STUDY OF COSMIC RAYS IN THE STRATOSPHERE IN THE VICINITY OF THE MAGNETIC EQUATOR

By S. N. VERNOV and A. V. MIRONOV

(Communicated by S. I. Wawilow, Member of the Academy, 28. II. 1939)

Studies of the dependence of the intensity of cosmic rays at high altitudes upon latitude afford the opportunity of defining the nature of primary cosmic radiation. Besides, on the ground of data on the absorption of cosmic rays in the atmosphere at different latitudes, it is possible to draw certain conclusions regarding the law of absorption of cosmic particles of definite energy.

In connexion with the fact that the study of cosmic rays at various altitudes of the stratosphere is most expediently carried out by means of sending up recording sets in balloon-sounds aud transmitting the readings over the radio, we developed a special method permitting to perform observational work of such kind (1).

The recording of cosmic rays was accomplished by means of a Geiger-Müller counter. The discharges in the counter were amplified by an amplifier and were made to actuate a relay. The latter switched on the anode circuit of the transmitter of a radio-sound of Prof. P. A. Molchanov system. The number of radio-signals thus corresponded to the number of discharges in the counter. The counting of the radio-signals was effected by several observers. Moreover, sound-recording of the radio-signals we accomplished with the help of the Shorin sound-recording apparatus (the shorinophone). Repeated listening to the records taken on the shorinophone showed satisfactory agreement with the results obtained by direct count by ear. Due to the availability of atmospheric discharges as well as to changes in the wave length of the transmitter, it was not possible to take automatic records during flight by means of an ondulator.

As a source of high potential for the counters, a high voltage portable battery of a special design with normal cells was used. With a view to keep working conditions constant, the temperature in the apparatus was maintained constant during the flight with a precision up to 5°. This was attained through the use of a chemical thermoregulator, in which heat was emitted, owing to a reaction between concentrated sulphuric acid and alkali. In our opinion, strict constancy of temperature in measurements of the latitude effect of cosmic rays in the stratosphere is an absolute necessity in order to keep constant the sensitivity of the counters during flight.

Besides radio-signals from counter, data on the altitude attained were also transmitted. For this purpose in the apparatus there were two counters of different size. With the help of a barograph switching over was effected, in result of which, alternately, either only the discharges in the smaller counter were transmitted, or else, the discharges in both the counters. Thus, the number of radio-signals sharply varied (about threefold)

over definite periods. These moments of sharp change in the number of radio-signals corresponded to the switchings over of the barograph. The barograph was taken with dissimilar length of contacts with a view to being able by the succession of contacts to judge of whether the balloon was ascending or descending.

In 1936, during the Erevan expedition ([2]), the existence of strongly pronounced latitude effect of cosmic rays in the stratosphere was found, proving that most cosmic rays possess a corpuscular nature. Simultaneously with this, the observed strong increase of the latitude effect with elevation showed that cosmic particles with an energy of $5 \cdot 10^9$ V suffer losses of energy which considerably exceed the losses on ionization. However, the cosmic rays, reaching the latitude of Erevan, still underwent a very strong absorption on passing the atmosphere, while most of the particles had to be endowed with an energy exceeding $8 \cdot 10^9$ V. This circumstance might be accounted for either by the availability of very heavy losses of energy for cosmic particles of greater energy, or by the existence of γ-rays within cosmic radiation.

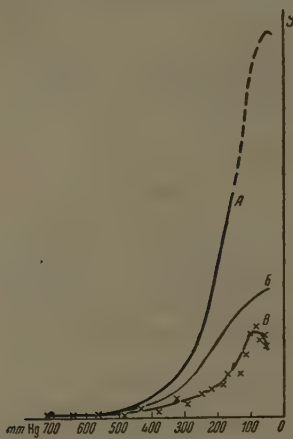

Fig. 1.

In order to settle this question we performed in 1937 and 1938 investigations of cosmic rays in the stratosphere near the magnetic equator. The flights of balloon with our apparatus were organized from on board the motoship «Sergo». Because of strong atmospheric discharges in the regions of the Red Sea and the Malacca Strait, most of the flights were effected in the Indian Ocean. The mean geomagnetic latitude, at which the flights were performed, was 5°. Five successful flights were carried out in equatorial regions. Besides this, two flights took place over the Black Sea.

The results obtained in the course of the flights are illustrated in Fig. 1 (curve *B*). Curves *A* and *Б* show the intensity of cosmic rays for the latitudes of Leningrad and Erevan. Comparison of the curves, showing the dependence of the intensity of cosmic rays upon altitude at different latitudes, shows that the intensity of cosmic rays in equatorial regions is four times less than for high latitudes. It was found possible on the basis of obtained data to evaluate the form of the spectrum of cosmic particles.

The area under the curve representing the dependence of the intensity of cosmic rays on altitude at the given latitude gives the energy, carried in by the cosmic particles.

If we make the simplifying assumption (it will be shown in the subsequent that this simplification does not involve any essential error) that at every latitude the magnetic field of the Earth admits cosmic particles with an energy exceeding the limiting one for the given latitude (in the vertical direction), which is determined according to the theory of Lemêtre and Vallarta, then the area comprised between curves *A* and *Б* will give us the energy carried by cosmic particles with energies from $1.5 \cdot 10^9$ V to $9 \cdot 10^9$ V, while the area between *Б* and *B* will show the energy of cosmic particles of $9—18 \cdot 10^9$ V. Lastly, the area under curve *B* gives the energy of particles with an energy exceeding $18 \cdot 10^9$ V.

Fig. 2 represents the spectrum of cosmic particles. Areas *I* and *II* of the rectangles are respectively equal to the areas between curves *A* and *Б*, *Б* and *B*. The smooth curve is plotted in such a way that the area for particles

with an energy exceeding $18 \cdot 10^9$V should correspond to the area under curve *B* in Fig. 1. In connexion with the fact that at high latitudes, according to Carmichael and Dymond's data ([3]), no latitude effect of cosmic rays is observed in the stratosphere, the spectrum is limited on the side of the lesser energies.

Since our measurements were accomplished with the help of a single counter capable of recording the cosmic rays which reached that counter from any direction, part of the cosmic particles possessing an energy below the limiting one for the given latitude would reach the set, while, at the same time, a certain portion of the particles with an energy above the limiting one would be deflected by the magnetic field. In conformity with Lemêtre and Vallarta's theory, the whole of the spectrum of primary cosmic particles may be divided into three regions. Part *III* gives the spectrum of the particles falling upon the top of the atmosphere in equatorial regions (5°), while part *II*—that of the particles causing the divergence of curves *B* and *B* in Fig. 1, etc. In Fig. 2 it is seen that the areas of part *I* and *II* are respectively almost equal to areas *I* and *II* of the rectangles. This shows that, with a spectrum of the aspect we have here, this spectrum can be fully well interpreted on the basis of the simplifying assumption that in a given latitude all particles with an energy exceeding the limiting one for the vertical direction are present. In developing the spectrum for primary cosmic particles, the assumption was made that there were no photons in cosmic radiation. As is shown in the following article, there is every reason to believe that the amount of γ-rays in the spectrum of primary cosmic rays should not be very great.

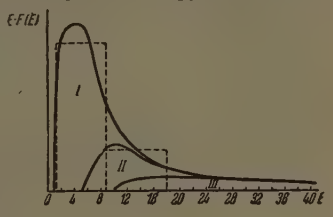

Fig. 2.

During our equatorial expedition there appeared in print a communication by Bowen, Millikan and Neher ([4]) dealing with the results of their study of the latitude effect of cosmic rays in the stratosphere. The evidence obtained by the aforesaid authors confirms the data of our measurements made in Erevan and agrees in the main with the results of our equatorial expedition. In comparison with the spectrum of primary cosmic particles found by Bowen, Millikan and Neher, the spectrum obtained by us shows the prevalence of soft particles, which fact may be brought in connexion with the existence of a longitudinal effect of cosmic rays in the stratosphere.

Measurements of the longitudinal effect at sea level show ([5]) that at the longitudes, under which our measurements were performed, the intensity of cosmic radiation is considerably lower than in the American hemisphere, where the investigations of Bowen, Millikan and Neher were partly conducted.

In closing we deem it our duty to express our sincere gratitude to S. I. Wawilow for a number of valuable suggestions and advices and for the organization of our expedition, as well as to N. L. Grigorov and J. I. Shaikhzamanov for participation in the work.

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4. III. 1939.

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PHYSICS

ANALYSIS OF THE LATITUDE EFFECT OF COSMIC RAYS IN THE STRATOSPHERE

By S. N. VERNOV

(Communicated by S. I. Wawilow, Member of the Academy, 28. II. 1939)

In the light of the data on the dependence of cosmic rays intensity upon altitude at various latitudes, a conclusion can be made as to the law of absorption of corpuscular rays of a definite energy.

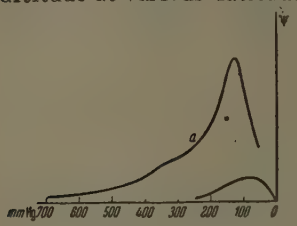

Fig. 1.

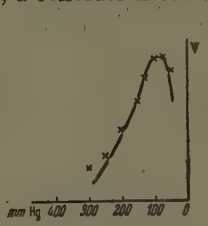

Fig. 2.

The difference between the curves represented in Fig. 1 of the foregoing article gives the absorption curve for particles, whose spectrum is shown in Fig. 2 of the same paper.

For comparison of the absorption curve obtained in this way with the conclusions of the cascade theory it will be necessary to determine the absorption curve for a parallel beam of cosmic rays.

This can be effected on the basis of the experimental data obtained with the help of Gross' formula (1):

$$2\pi i_0 \psi(x) = I_{\frac{\pi}{2}}(x) - x\frac{dI_{\frac{\pi}{2}}(x)}{dx},$$

where i_0 is the intensity of cosmic radiation within a unit of solid angle at the top of the atmosphere, $\psi(x)$ is the intensity of a vertical beam of cosmic rays and $I_{\frac{\pi}{2}}(x)$—the intensity of cosmic rays reaching the apparatus from all directions $\left(\text{from } 0 \text{ to } \frac{\pi}{2}\right)$ corresponding to absorbing layer x.

However, in connexion with the existence of a sharply marked latitude effect in the stratosphere, cosmic radiation at the top of the atmosphere will fail to be isotropic in equatorial regions. Owing to this fact, the possibility of the application of Gross' formula leads here to doubts, since this formula was deduced on the assumption that radiation at the top atmosphere is isotropic.

The possible error in the calculation by Gross' formula, using the obtained data near the magnetic equator, may be determined in the follow-

ing way. The intensity of cosmic rays, reaching the apparatus at angles from 0 to ϑ vertically (I_ϑ) is determined according to formula

$$I_\vartheta(x) = 2\pi i_0 \int_0^\vartheta \psi\left(\frac{x}{\cos\vartheta}\right) \sin\vartheta\, d\vartheta. \qquad (1)$$

Differentiating equation (1) it is easy to show that, if the dependence of the intensity of cosmic radiation reaching the apparatus at angles from 0 to ϑ upon altitude be known, then $\psi(x)$ is obtainable from equation

$$2\pi i_0 \left[\psi(x) - \psi\left(\frac{x}{\cos\vartheta}\right)\right] = I_\vartheta(x) - x\frac{dI_\vartheta}{dx}. \qquad (2)$$

If the difference of I_ϑ from $I_{\frac{\pi}{\text{II}}}$ is neglected and the approximate value of $\psi_{\text{appr.}}(x)$ be calculated according to Grosss' formula substituting I_ϑ for $I_{\frac{\pi}{2}}$, then this value differs from the true value; yet, the involved error may be evaluated from equality

$$\psi_{\text{appr.}}(x) = \psi(x) - \psi\left(\frac{x}{\cos\vartheta}\right)\cos\vartheta.$$

In a similar way, if the dependence of the intensity of cosmic radiation reaching the apparatus at angles from ϑ to $\frac{\pi}{2}$ ($I_1 = I_{\frac{\pi}{2}} - I_\vartheta$) upon altitude be known, then $\psi(x)$ is determined from the equality

$$2\pi i_0 \psi\left(\frac{x}{\cos\vartheta}\right)\cos\vartheta = I_1(x) - x\frac{dI_1}{dx}.$$

Knowing the dependence upon altitude of cosmic radiation consisting of various components, one of which reaches the apparatus from all directions, while the other one within a cone from 0 to ϑ_1, etc., it will be possible to construct analogous equalities.

Determining on the basis of the latitude effect the angular distribution of cosmic rays at the top of the atmosphere, it will be possible to divide the cosmic radiation into a number of such components.

By the action of the earth-magnetic field the positively charged cosmic ray particles (for computation of the maximum error it will be assumed that all the particles are positive) reach the apparatus only within a cone with an axis directed westward. In view of the fact that the distribution of particles along the azimuth is unessential for the determination of the error in our conversion to the terms of the Gross formula, we can reduce the problem to the one considered above by determining the angular distribution of the primary particles falling on the top of the atmosphere, doing this on the basis of the knowledge of the spectrum of primary particles.

Calculations performed in the above-indicated way have shown that the error committed in the computation of the intensity of a vertical beam according to Gross' formula is inconsiderable. Up to an altitude of 16 km, this error ought not to constitute more than 6%. For greater altitudes this error increases slightly and at altitudes over 20 km the excess of the calculated values over the true ones may amount to 20%. These deviations are, in the main, caused by the existence in cosmic radiation at equatorial regions of particles with an energy lower than $18 \cdot 10^9$.

The calculations performed above show that with the indicated error it is possible to determine the intensity of the vertical beam by Gross' formula. The dependence of the intensity of a vertical beam of cosmic rays upon altitude for a latitude of 5° obtained in this way is represented in Fig. 1 (curve *a*).

For comparison of the absorption of cosmic particles with the conclusions of Serber's [2] cascade theory, data are represented in Fig. 2 (denoted by crosses) giving the difference between the absorption curves for a vertical beam of cosmic rays at a latitude of 50° (according to Pfotzer's data) and at a latidude of 4°.

For comparison, a curve is plotted which has been calculated on the basis of Serber's theory for cosmic particles, whose spectrum is known from the data given in our preceding article. For great altitudes there is satisfactory agreement, while at low altitudes some divergence is to be observed, which must be connected with the presence of mezotrons and electrons formed in result of the disintegration of mezotrons.

The existence of a strongly expressed latitudinal effect in the stratosphere shows that a considerable amount of the energy supplied by cosmic rays is carried by electrons or positrons.

Since, owing to the action of the magnetic field of the earth, the intensity of cosmic radiation at the equator is four times less than at great latitudes, we may consider that at least 75% of the total energy of cosmic radiation is carried by corpuscular rays. However, on the basis of the shape of the absorption curve in equatorial regions, the deduction can be made that only an insignificant portion of the total energy of cosmic radiation can fall on the top of the atmosphere in the form of γ-rays produced by cosmic electrons, previous to their entering the magnetic field of the Earth.

The mean energy of γ-rays produced by primary electrons should necessarily be considerably lower than the energy of the electrons admitted by the magnetic field of the Earth close to the equator. These rays should, therefore, be strongly absorbed at passing through the atmosphere.

Fig. 1 (curve δ) represents the absorption of γ-rays (as well as the secondary rays produced by these γ-rays in the atmosphere) created by primary electrons, having a known spectrum at passing through a layer of matter equivalent to 1.5 cm of water. Calculations have been made on the basis of Serber's theory.

The difference between curves a and δ in Fig. 1 gives the absorption curve for electrons. It will be seen from Fig. 1 that the existence of a large amount of γ-rays is outside the limits of possibility, since should this be the case, a very strong rise of the curve for electrons from $P=50$ mm Hg up to 140 mm Hg would then occur, which stands in contradiction with the deductions of the cascade theory. As shown by Serber's curves for small thickness of the absorbing layer, a linear increase takes place for particles with not very high energy.

Thus, the conclusion can be drawn that cosmic rays before entering the magnetic field of the Earth are unable to pass a layer of matter exceeding 1.5 g and, moreover, that less than 3% of the total energy of cosmic radiation fall at the top of the atmosphere is carried by secondary γ-radiation.

In closing I deem it a pleasant duty to express my deep gratitude to Prof. D. V. Skobeltzyn for valuable suggestions and advices.

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PHYSICS

HALL-EFFECT IN AN $AuCu_3$ ALLOY IN AND OUT OF THE ORDERLY CONDITION

By A. KOMAR and S. SIDOROV

(Communicated by S. I. Wawilow, Member of the Academy, 5. III. 1939)

It is well known that the planes separating a K-surface into separate Brillouin zones will, in the case of a cubic lattice, run parallel to the planes which, according to Bragg, are liable to reflect X-rays ([1],[2]).

It is also known ([1]) that with electrons but inconsiderably filling the zones, corresponding to high energy levels, the Hall-effect can very well be computed, according to classical formulae. Now, if the structure of the metal has undergone change and if new reflecting surfaces have appeared, the form of the zones is also liable to change. It may happen that a zone corresponding to a high level of energy will be greatly filled and in this case the Hall-effect will change its sign, wherefore it, of course, can no longer be simply calculated according to classical formulae. Usually the calculation of such an anomalous effect is performed in the classical way, but making use of the conception of «holes» carrying a positive charge. Under practical conditions, it is of course possible to envisage in metals and especially in alloys combinations of the «hole» and the «normal» Hall-effect.

On the basis of these general considerations, one would be likely to expect that the Hall-effect of intermetallic compounds giving new reflexion planes should be anomalous in respect of magnitude and sign, as compared with the Hall-effect of an ordinary alloy in the non-orderly state, and, indeed, a number of investigations of Evans and co-workers ([3]) point to this view being entirely correct. It appeared to us that the afore-stated considerations could be confirmed with particular clarity for the case of the Hall-effect of phases passing into the ordered condition. It is known ([4]) that the Hall-constant for binary solid solutions is a monotonous function of concentration. When an alloy passes into the ordered state, new reflecting surfaces make their appearance, the form of the zones becomes changed and there is great likelihood that the Hall-effect will become materially changed.

In order to verify the said hypothesis, we measured the Hall-constant for the $AuCu_3$ alloy in and out orderly state. The alloy in question suited our purpose very well. First, it is an alloy which has been thoroughly studied with the help of X-rays, and secondly, on passing into the state of order it does not undergo any change in symmetry, while the lattice constant changes but very little (0.16%). The effect of changes in symmetry and lattice constant, thus, does not enter the picture. Thirdly, the Hall-effect for the pure components stands in agreement with the classical theory.

The sequence of changes in the alloy was as follows: not in orderly state →in orderly state, in orderly state→not in orderly state, not in orderly state→ in orderly state. Measurements were effected for all states on the same two plates. The thickness of the plates was slightly uneven, so that we do not insist on the absolute values of the constants, being desirous of stressing the fact of the change only. The temperature of the transition from the orderly to the non-orderly state was for our alloy∼382°C. We obtained the non-orderly state by hardening from a temperature of 600°C and the orderly state by annealing at a temperature lying below the Curie point.

Results of the 1 st Series of Measurements

Quenching from 600°
Non-ordered state: 1st plate $R = -597 \cdot 10^{-6}$ C. G. S. M.
» » » 2nd » $R = -581 \cdot 10^{-6}$ C. G. S. M.
Double-step annealing at 370° — 24 hours and then at 360° — 29 hours
Ordered state: 1st plate $R = -191 \cdot 10^{-6}$ C. G. S. M.
» » 2nd » $R = -171 \cdot 10^{-6}$ C. G. S. M.

Results of the 2 nd Series of Measurements

Quenching from 600°
Non-ordered state: 1st plate $R = -604 \cdot 10^{-6}$ C. G. S. M.
» » » 2nd » $R = -577 \cdot 10^{-6}$ C. G. S. M.
Double-step annealing at 370° — 24 hours and then at 360° — 46 hours
Ordered state: 1st plate $R = -230 \cdot 10^{-6}$ C. G. S. M.
» » 2nd » $R = -210 \cdot 10^{-6}$ C. G. S. M.

Thus, we ascertained an almost threefold change in the Hall-constant in course of the passage of the alloy from disorderly to the ordered state.

It is necessary to note that at 360° full order has not yet set in in the alloy; on lowering the temperature, there are all reasons to expect the onset of more orderly conditions and greater changes in the Hall-constant. As a matter of fact the Hall-constant of the specimens annealed at 325° equalled $65 \cdot 10^6$ C. G. S. M.

Thus, we can presume the dependence of the Hall-constant on the extent of the attainment of orderliness in the alloy. We have undertaken further work along the lines of settling this problem.

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GEOPHYSICS

A FEW WORDS ON THE REDUCTION OF GRAVITY

By N. MIHAL

(Communicated by S. I. Wawilow, Member of the Academy, 5. III. 1939)

First of all let us introduce some definitions and notations. Let S be a level ellipsoid, whose centre is combined with the centre of the Earth's mass, while its smaller axis is directed along the rotation axis of the Earth; S' is the geoid, S''—the physical surface of the Earth; T—the disturbing potential, i. e. the difference between the potentials of the Earth's gravity and the level ellipsoid; A, B, C... are points of the physical surface of the Earth having respectively elevations H_a, H_b, H_c, ... above the geoid. The masses producting T, which are to be found both inside and outside the geoid, be denoted as disturbing masses.

Let us prove the irrelevancy of the frequently accepted interpretation of H. Faye's anomaly usually stated under two versions. In the one of these versions the continents become «immersed» into the geoid down to different depths H_a, H_b, H_c, ..., whereupon correction $\frac{2G_0}{R} H_i$ $(i=a, b, c,...)$ is introduced into the gravity force, so that:

$$\left.\begin{aligned} \left(\frac{dT_1}{d\rho}\right)_{S'} &= G_a + \frac{2G_0}{R}(H_a + N_a) - \gamma_{S'}, \\ \left(\frac{dT_2}{d\rho}\right)_{S'} &= G_b + \frac{2G_0}{R}(H_b + N_b) - \gamma_{S'} \\ &\cdots\cdots\cdots\cdots \\ &\cdots\cdots\cdots\cdots \end{aligned}\right\} \qquad (1)$$

where $\gamma_{S'}$ is normal gravity upon S', brought about by the mass of the level ellipsoid; N is the distance between the level ellipsoid and the geoid. But this interpretation is altogether erroneous on account of two reasons: 1)correction $\frac{2G_0}{R} H_i$ is lawful only for normal Earth (of the level ellipsoid); 2) T, T_1, T_2 ... have nothing in common with one another, as they correspond to different schemes of the Earth, so that at an outer point $T \neq T_1 \neq T_2 \neq \ldots$; consequently, the joint use of anomalies (1) is impossible. The first reason may be eliminated, if all disturbing masses be immersed together with point A, but the second reason remains in force.

The second variant of the interpretation of Faye's anomaly differs from the first one only by the fact that instead of the immersion of continents, their condensation on the surface of the geoid is assumed. This interpretation, as will easily be seen, is fallacious for the very same reasons.

Thus, the usual interpretation of Faye's anomaly leads to incorrect, illusive conceptions. Faye's anomaly should be interpreted in a different manner, namely:

Normal gravity at point A, lying on the physical surface of the Earth, is equal to:

$$\gamma_S - \frac{2\gamma_0}{R}(H+N).$$

Consequently, the difference between the gravity of the Earth and normal gravity at point A is determined by formula:

$$G + \frac{2\gamma_0}{R}(H+N) - \gamma_S. \qquad (2)$$

If we discard correction $\frac{2\gamma_0}{R}N$, we shall obtain Faye's anomaly, i. e. obtain the difference between the gravity of the Earth at point A and the normal gravity at a point lying at distance N above or below point A, according to the sign of N.

In making use of G. Stokes' formulae, anomalies $G + \frac{2G_e}{R}H - \gamma$ (1) will be altogether suitable. But, on purely geological grounds, it will be necessary to have recourse to anomalies (2), bearing in mind that they refer to points of the physical surface of the Earth.

In conclusion let us note that anomalies (2), contrary to the usual interpretation, belong to one and the same scheme of the Earth, and accordingly their joint application is quite possible.

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GEOPHYSICS

DETERMINATION OF ORIENTATION OF THE ELLIPSOID OF EARTH'S INERTIA

By N. MIHAL

(Communicated by S. I. Wawilow, Member of the Academy, 5.III.1939)

Let A, B, C, F, G, H be moments of inertia and products of the Earth's inertia referred to the system of co-ordinates $Oxyz$. The equation of the ellipsoid of inertia has in the given system the appearance:

$$Ax^2+By^2+Cz^2-2Fzy-2Gzx-2Hxy=1. \tag{1}$$

The coefficients with squares of co-ordinates of the transformed quadratic form (1) to the main axes are the roots of the equation:

$$\begin{vmatrix} A-\lambda, & -H, & -G \\ -H, & B-\lambda, & -F \\ -G, & -F, & C-\lambda \end{vmatrix}=0. \tag{2}$$

Or, opening the determinant

$$-\lambda^3+m\lambda^2+n\lambda+q=0, \tag{3}$$

where

$$\left.\begin{aligned} m&=A+B+C, \\ n&=F^2+G^2+H^2-(AC+BC+AB), \\ q&=ABC-AF^2-CH^2-BG^2-2FGH. \end{aligned}\right\} \tag{3'}$$

By substituting

$$\lambda=y+\frac{m}{3} \tag{4}$$

equation (3) will be reduced to the form:

$$y^3+ay+b=0, \tag{5}$$

where

$$a=-\left(n+\frac{1}{3}m^2\right); \quad b=-\left(\frac{2}{27}m^3+\frac{nm}{3}+q\right). \tag{5'}$$

Upon simple transformations connected with the application of formulae (3′) and (5′) we obtain:

$$\left.\begin{aligned} a&=-(F^2+G^2+H^2)-\tfrac{1}{3}[(C-A)^2+(A-B)(C-B)], \\ b&=-\tfrac{2}{27}[(A-B)-(B-C)]^3-\tfrac{1}{3}[(A-B)-(B-C)][F^2+G^2+ \\ &\quad +H^2+(A-B)(B-C)]+(A-B)F^2+(C-B)H^2+2FGH. \end{aligned}\right\} \tag{6}$$

The differences between A, B and C, as well as the coefficients F, G, H, are determined by gravimetrical data (1) (the formulae deter-

mining F, G, H are given at the end of the present paper). Accordingly, the roots of equation (5) and therefore also the differences of the roots of equation (3), i. e. the differences of the principal moments of the Earth's inertia, can be determined by gravity observations.

Let us now proceed to investigations on the determination of the orientation of the ellipsoid of the Earth's inertia.

Let α, β, γ be the directrix cosines (in the $Oxyz$ system) of any of the axes of the ellipsoid of inertia. Then, as is known,

$$\begin{array}{lll} (A-\lambda)\alpha & -H\beta & -G\gamma=0 \\ -H\alpha+(B-\lambda)\beta & & -F\gamma=0 \\ -G\alpha & -F\beta & +(C-\lambda)\gamma=0 \end{array}$$

or, taking into consideration (4) and (3′),

$$\left.\begin{array}{lll} \left\{\frac{1}{3}[(A-B)+(A-C)]-y\right\}\alpha-H\beta & & -G\gamma=0 \\ -H\alpha+\left\{\frac{1}{3}[(B-A)+(B-C)]-y\right\}\beta & & -F\gamma=0 \\ -G\alpha & -F\beta+\left\{\frac{1}{3}[(C-A)+(C-B)]-y\right\}\gamma & =0 \end{array}\right\} \quad (7)$$

Thus, the system of equations (7) together with equation $\alpha^2+\beta^2+\gamma^2=1$ gives the directrix cosines of the axes of the ellipsoid of the Earth's inertia, provided the differences between A, B and C and coefficients F, G, H are known. By the method expounded in our article [1] it will be possible to obtain the necessary formulae for determining F, G, H. We shall write them out without deduction

$$H=\frac{\Delta J_3}{4\pi f}, \quad (8)$$

where

$$\Delta J_3=a^2\int_\Sigma(g-\gamma)N(\theta,\lambda)\,d\sigma+a^2\left(2+\frac{9}{14}\,q-\frac{5}{7}\,\alpha\right)\int_\Sigma\Delta g\sin^2\theta\sin 2\lambda\,d\sigma-$$

$$-a^2\left(\frac{5}{3}\,q-\frac{4}{3}\,\alpha\right)\int_\Sigma\Delta g\sin^4\theta\sin 2\lambda\,d\sigma,\quad N(\theta,\lambda)=\left(\frac{1}{2}-\alpha\cos^2\theta\right)\sin^2\theta\sin 2\lambda.$$

$$G=\frac{\Delta J_4}{4\pi f}, \quad (9)$$

where

$$\Delta J_4=a^2\int_\Sigma(g-\gamma)M(\theta,\lambda)\,d\sigma+a^2\left(2-\frac{1}{14}\,q-\frac{1}{7}\,\alpha\right)\int_\Sigma\Delta g\sin^2\theta\sin 2\lambda\,d\sigma+$$

$$+a^2\left(\frac{5}{3}\,q-\frac{4}{3}\,\alpha\right)\int_\Sigma\Delta g\sin^4\theta\sin 2\lambda\sin^2\lambda\,d\sigma,$$

$$M(\theta,\lambda)=\left(\frac{1}{2}-\alpha\sin^2\theta\sin^2\lambda\right)\sin^2\theta\sin 2\lambda.$$

In this last formula the following notations are adopted: the polar axis is directed along axis Oy; θ is the polar distance of point A having flowing co-ordinate; λ is the angle between the plane of the equator and the plane of the meridian counted off counter-clockwise from the positive axis Ox. To determine F formula (9) is used, provided the polar axis be directed along axis Ox. In (8) and (9) due account is taken of the fact that F', G' and H' for a level ellipsoid are equal to zero.

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GEOPHYSICS

A GRAPHICAL METHOD OF TREATMENT OF SEISMOGRAMS OF PLUTONIC EARTHQUAKES

By A. A. TRESKOV

(Communicated by S. I. Wawilow, Member of the Academy, 5. III. 1939)

The tables of Gutenberg and Richter (1) contain the travel times of all seismic waves as functions of the epicentral distance Δ and the depth of focus h, while the usual method of determining h and Δ is based not upon the assemblage of phases stated, but only on differences $pP—P$, $sS—S$ and $sP—P$, the complete utilization of the tables thus being limited.

In this connexion the problem arises of devising a method of determining the depth of focus and the epicentral distance, which should utilize all the phases and be directly demonstrative, thus facilitating the rational application of Gutenberg's tables and securing greater reliability and correctness of interpretation of phases recorded in seismograms.

In the writer's opinion, the above requests should be satisfied by a graphical method advanced by him and based upon the following principles.

On the ground of tables of Gutenberg and Richter, a curve is drawn in co-ordinates of h and Δ, based upon the difference $f—P$ of the times of arrival of the waves of the given phase and of longitudinal waves. The co-ordinates of the points of the curve determine the total of combinations of values of h and Δ, corresponding to the given difference $f—P$, in conformity with the tables of Gutenberg and Richter. If curves are constructed for all phases stated, then, in the case of perfectly accurate definition of the times of phases and of ideal precision of tables of Gutenberg and Richter, all the curves drawn must intersect at one and the same point, of which the co-ordinates will yield the exact values of the epicentral distance Δ and of the depth of focus h.

Of course, in practice, the intersection of all the curves does not occur at one point, but, for correct interpretation of phases whose moments are established with sufficient accuracy, the curves intersect within a limited area, which we denominate the n u c l e u s of the diagram and whose dimensions give a clear idea of the limits of precision of the solution.

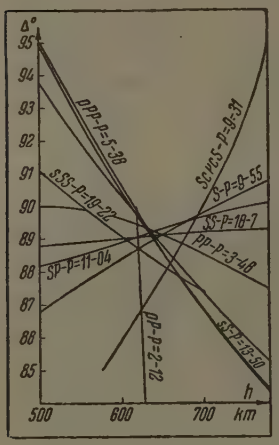
The graphical treatment requires first of all the supposition as to the nature of several phases distinctly recorded in the seismogram. In the case of correct interpretation of phases, the curves of the diagram are intersecting within the limits of a given area (nucleus); should the curves run unconformably, this would point to the fact that our assumption as to the nature of phases was erroneous. Thus, from the very beginning we are enabled to control the correctness of the resolution of the problem.

When, utilizing several phases, the nucleus of the diagram is identified, it is easy to show the approximate values of the depth of focus and of the

epicentral distance, while the said evidence together with the tables of differences $f-P$ (or $f-P'$) for all the phases, computed from the tables of Gutenberg and Richter, allow easily to recognize the nature of the remaining phases and to plot respective curves on the diagram.

Curves which are in poor agreement with most of the other curves, as well as single points of intersection beyond the limits of the nucleus, are not to be taken into consideration on treatment, these anomalous elements of the diagram being possibly called forth by an inaccuracy in determining the times of phases and by the small angle of relative slope of the curves intersecting.

Moscow



In order to determine the depth of focus and the epicentral distance, the writer makes use of the following scheme. The co-ordinates of h and Δ are determined of the points of intersection lying within the limits of the nucleus, and for the final values of the focus depth and of the epicentral distance are taken arithmetical means of the co-ordinates of the totality of the above points, the point of intersection of three curves being taken equivalent to three points of simple intersections, while the point of intersection of four curves is considered as equivalent to six simple points (numbers of combinations by two of three and four elements).

As an example of diagram illustrating the graphical method of the author, may serve the annexed graph based upon the seismogram of the earthquake of August 11, 1937, as recorded by the station Moscow.

The results computed in the table and obtained for the given case by the graphical method from the data of the stations Sydney (Riverview), Bombay, Wladivostok, Irkutsk, Sverdlovsk, Moscow, Pulkovo and Strasburg, allow to draw the following conclusions:

Data for Earthquake of August 11, 1937; $\varphi_e = 7°S$, $\lambda_e = 116°E$, $h = 636$ km

Stations	Co-ordinates of stations		Depth of focus h	Distance from diagram Δ	True distance Δ_0	$\Delta_0 - \Delta$	Time of longitudinal waves	Travel time of waves P after Gutenberg	Time of focus
	φ	λ							
Sydney .	33°50'S	151°10'E	(640)	(43°.6)	42°6	(−1°0)	$1'2^m59^s(i\,P_2)$	7^m8^s	$0^h55^m51^s$
Bombay .	18°54'N	72°49'E	627	49.6	49.8	+0.2	1—3—55	8— 1	0—55—54
Wladivostok . .	43° 7'N	131°54'E	640	51.4	52.2	+0.8	1—3—58?	8—17	0—55—41?
Irkutsk .	52°16'N	104°19'E	640	60.9	60.1	—0.8	—	—	—
Sverdlovsk .	56°50'N	60°38'E	584	78.5	78.0	—0.5	1—6—48	10—59	0—55—49
Moscow .	55°44'N	37°35'E	630	89.2	89.3	+0.1	1—7—51(iP)	11—55	0—55—56
Pulkovo .	59°46'N	30°19'E	616	92.0	93.9	+1.9	1—8—9	12—16	0—55—53
Strasburg	48°35'N	7°36'E	626	106.2	107.3	+1.1	1—9—12	13—15	0—55—57

1) the graphical method of Treskov permits to determine the depth of focus to an accuracy of 10—20 km from the data of one station;

2) the possible error of epicentral distance is of the order of 1°;

3) the time values of the focus, determined for various stations from the tables of Gutenberg and Richter, are not inferior in agreement to corresponding data for ordinary earthquakes, which points to a sufficient accuracy of Gutenberg's tables with regard to travel times of longitudinal waves.

In conclusion it should be noted that the graphical method can be utilized for the study of ordinary earthquakes. For a superficial quake, the curves of the diagram must (theoretically) intersect at the axis Δ ($h=0$) and the displacements of the nucleus of the diagram from the said axis may, if but approximately, clear up the question of the real depth of focus in different seismic regions.

When treating a normal earthquake by the graphical method, the epicentral distance will be determined from the total of phases and not from the difference $s-P$; the data of the graphical treatment should also be utilized for corrections of the tables of travel times; however, investigations of such kind will not be of real use, unless the tables of Gutenberg and Richter will be sufficiently verified to become fully reliable with respect to a series of main phases.

Seismic. Station Irkutsk.

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CHEMICAL TECHNOLOGY

A CALCINED MULLITE REFRACTORY

By I. I. KITAIGORODSKI and T. N. KESHISHJAN

(Communicated by I. V. Grebenshikov, Member of the Academy, 22. II. 1939)

The glass industry has of late adopted the use of an electrically-cast mullite refractory of the Corhart type, the production of which involves the use of electric melting of a highly aluminaceous charge with subsequent crystallization and cooling of the bars cast.

Due to the complexity of the manufacturing process, the cost of this refractory exceeds several times the outlays, needed for the production of materials through the ordinary methods used in the ceramic industry. However, the refractory in question is widely used, since it possesses two properties which make it highly valuable for the glass industry, to wit:

a) high refractoriness;

b) high glass resistance.

In 1936, the Laboratory of glass technology of the Mendeleyev Chemical Technology Institute set itself the task of obtaining by means of calcining a refractory which, with regard to its properties, would in no respect fall short of electrocast mullite of the Corhart type. For this purpose the system SiO_2—Al_2O_3—MgO, whose equilibrium diagrams provided a basis for a study of the most desirable compositions, was selected. From among all the mixtures tried, a composition of 72% Al_2O_3, 25% SiO_2 and 3% MgO was found to answer the requirements in the best way.

The following raw materials, having a composition as stated below, were taken:

	SiO_2	Al_2O_3	Fe_2O_3	CaO	MgO	Loss on ignition
Alumina	0.95	89.95	0.09	—	—	9.34
Kaolin	43.91	39.16	0.48	1.48	0.61	14.25
Magnesium oxide . .	Chemically pure					

The charge estimated to yield 100 g biscuit was moistened with 20—25 cm^3 water.

Standard cylinders with a diameter of 38 and a height of 45—48.5 mm were next made by ramming in metallic cylindrical molds. The drying of the cylinders was effected during 2 days at a temperature of 110°. The calcining continued 24 hours and was carried out in a kerosene furnace, the holding at 1500° lasting for 5—6 hours. After the first calcining the investigation of samples showed the following:

Igneous shrinkage	10.6%
Voluminar weight	2.02 g/cm^3
Water absorption	16.03%
Apparent porosity	36.0%

The second calcination of the same samples was effected in krypton furnaces. The calcination continued for 5 hours. The maximum temperature of calcination was 1725° and it was held for 30 min. The second calcination gave an additional 10% shrinkage.

Thin sections of the specimens taken after the second calcination showed the following. The specimens contained large single crystals and bundles of mullite lying in the midst of a finely crystalline mullite mass. Besides mullite, the specimen was also found to contain small grains of corundum, which were perceptible both in the fine-grained ground-mass and in the big crystals.

The following properties of the specimens were determined: refractoriness, voluminar weight, water absorption, apparent porosity, specific gravity, content of mullite and corundum in biscuit (treated with fluoric acid), glass resistance, and thermal strength (tested by immersion into water of the test samples heated up to 850°).

We deem it wise to enlarge upon some details of the method used for the determination of glass resistance.

For a comparative determination of the glass resistance of the new kind of mullite obtained, we manufactured small bricks out of this material, out of cast mullite biscuit and out of ordinary pot material. These small bricks were next subjected to special treatment. i. e. were boiled in baryte crown having the following chemical composition:

SiO_2 — 32.20
B_2O_3 — 13.20
BaO — 45.90
ZnO — 3.50
Al_2O_3 — 3.15
As_2O_3 — 1.60

The holding lasted for 6 hours. During the whole time of the holding, the temperature was kept at 1300°.

After boiling in this glass, the bricks were cleaned from any sand that might have adhered to them by concentrated muriatic acid of double dilution (sp. gr. 1.19).

The loss in weight of the bricks showed that the attack of the glass on them was negligible. The results of the testing are to be found in the table.

Material	Weight of sample, g	Surface, cm^2	Loss in weight, g	% loss in weight	Loss in weight, g/cm^2 hour
Ordinary pot material	2.5058	7.26	0.7288	29·0	0.016
Cast mullite	3.7610	7.26	0.4760	12.7	0.011
Material investigated (calcined mullite)	3.4550	7.26	0.3700	10.6	0.008

This table helps to visualize the high glass resistance of the material obtained.

Conclusions

1. In result of the use of a charge containing definite proportions of SiO_2—Al_2O_3—MgO system a highly resistant refractory has been evolved.

2. The initial raw material consisted of kaolin, pure aluminium oxide and magnesium oxide.

3. The specimens were produced by ramming.

4. There were two calcinations; one at 1500° with holding for 5—6 hours at this temperature; the second one at 1700—1725° with holding for 30 min. at this temperature.

5. The refractory thus obtained in 1936 by the Laboratory of the Mendeleyev Chemical Technological Institute is characterized by the following data:

Contents of mullite and corundum in biscuit	81.7%
Refractoriness	1800°
Voluminar weight	2.68—2.69
Water absorption	0.5—3.4%
Apparent porosity	1.2—9.0%
Specific gravity	3.186—3.189
True porosity	15—16%
Glass resistance (loss in weight)	0.008 g/cm² hr
Thermal strength	11—36 thermal changes

6. The results obtained under laboratory conditions warrant the expediency of starting a check of the properties of the new refractory on an industrial scale.

Mendeleyev Institute of Chemical Technology.
Laboratory of Glass Technology.

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BIOCHEMISTRY

PARTICULARS IN ENZYMATIC SYSTEM OF VARIOUS PEAS

By E. V. ARCIKHOVSKAJA and N. S. SPIRIDONOVA

(Communicated by A. N. Bach, Member of the Academy, 28. II. 1939)

The regularity established by Rubin ([2],[3]) in his experimenting with the plants which store up matter chiefly in the form of sugars is that their ability to ripen more or less late is related to the synthetic activity of the sucrose and to how much the variety concerned is able to increase in sugar.

While working together with Lutikova ([6]) and Arcikhovskaja ([4],[5]) he found too that oxidizing processes direct the action of sucrose toward synthesis. The relationship between the activity and directional properties of enzymes, on the one hand, and the physiological as well as practically valuable characters of plants, on the other, has as yet been studied in our laboratory on those plants only which lay up sugars as reserve material. Now, however, we have started experiments with the plants which besides the carbohydrates store considerable amounts of other substances, protein in particular.

The present communication is dealing with our first attempts to inquire into the rôle of enzymes in peas. The experiment was run with two different kinds of pea: grain-pea and sugar-pea. Three varieties of the former (Late 1632, Victoria Mandorf—intermediate with regard to the time of ripening, and Early green) and two varieties of the latter [Albanian late, Lightning (Molnia) early] have been tested.

Seeds for planting grain-pea varieties were grown in two localities, strikingly different as regards climatic conditions—Pushkin and Middle Asia. No definite relation between the origin of seed and the biochemical processes in plants have we detected as yet. All the data on grain-pea tabulated below are averages from what has been recorded in plants originating from different geographical points. Since in a pea seed a great amount of nutrient matter is represented by proteins, we were endeavouring to find a relation between the «earliness» of the plant and the synthetic activity of its protease [the latter being determined according to Kursanov ([1])]. The data in Table 1 are the results of two tests.

From the figures referred to it may be seen that the synthetic action of the protease is higher in the late varieties, while it is zero in the early ones. The latter circumstance should be ascribed to the fact that in our test we had to deal with leaves and young pods, not with ripening grain, where the protein is chiefly laid up. The processes of fermentative synthesis of protein, which doubtlessly take place in those organs too, are probably so slow in early varieties, as to escape the record of 15 min. exposure used in the experiment.

Table 1

Synthetic Activity of Protease in Peas
(in mg of protein per 10 g of dry matter synthesized during 15 min.)

Pea type	Variety	Synthesis of proteins	
		in leaf	in pod
Grain-pea	1632	5.1	4.0
»	Victoria Mandorf	1.5	2.2
»	Early green . . .	0	0
Sugar-pea	Albanian	—	2.6
»	Lightning	—	0

The directive action of the protease in peas appears, thus, to be governed by the same regularity, as has been established for sucrose in sugar storing plants: the later the variety, the more intensively proceeds the synthesis. Such a behaviour of the protease falls in line with the content of protein in the grain of the varieties tested (Table 2).

Table 2

Pea type	Variety	% of protein	Ratio protein N / total N
Grain-pea	1632	25.4	85
»	Victoria Mandorf	24.2	86
»	Early green . . .	22.9	84
Sugar-pea	Albanian	28.8	80
»	Lightning	21.5	75

If the differences between the varieties as to the content of protein are not very great, especially among grain-peas, an unmistakable tendency may, nevertheless, be observed within either of the two types: the later the variety and the stronger the synthetic action of the protease, the higher is the content of protein in the grain.

The ratio of protein nitrogen to total nitrogen is nearly the same in the three grain-peas tested. However, within the sugar-pea type the late variety has this ratio markedly enhanced as compared to the early one.

Similar relations have been observed by Tovarnitsky in soya. The varieties heavily stored with proteins showed a smaller ratio of water-soluble protein and a weaker activity of the urease(7).

This reminds of the ratio of sucrose to monose which is typical for sugar-storing plants. Rubin (2,3) has shown that among the plants, where carbohydrates predominate in metabolism, the late varieties are higher in sugars in general and in sugars with complex molecule in particular. It would appear that in both cases we meet with the same phenomenon which is bound up with the synthetic action of sucrose or protease respectively.

We failed in establishing a relation between the oxidizing action of ascorbinase and the earliness of grain-peas. No observations on glutation were made in the course of the experiment, though this would be an interesting point to clear up, as it may be inferred from the works of Grassman(8), Waldschmidt-Leitz and Mothes (9).

The significance of ascorbinase activity becomes clear enough when comparing it to the directive action of sucrose. As seen from Table 3, the

Table 3

Directive Action of Sucrose (in mg of glucose per 10 g of dry matter, stored in 3 hours) and Oxidizing Activity of Ascorbinase (in mg of ascorbic acid per 10 g of dry matter per 15 min.)

Pea type	Variety	Ascorbic acid in leaves	Carbohydrates in grain (% of abs. dry weight)		Directive action of sucrose					
					Leaves			Pods		
			soluble sugars	starch	hydrolysis	synthesis	synth. hydrol.	hydrolysis	synthesis	synth. hydrol.
Grain-pea	1632	14.6	5.18	60.01	10.6	61.8	5.8	2.9	17.8	6.3
Grain-pea	Victoria Mandorf	15.6	4.66	62.47	4.5	58.6	13.0	6.0	50.4	8.4
Grain-pea	Early green .	17.7	6.05	61.57	0	81.1	—	—	—	—
Sugar-pea	Albanian . .	17.2	8.63	49.31	8.3	81.1	9.8	6.0	21.6	3.6
Sugar-pea	Lightning . .	15.3	4.56	56.7	32.7	51.5	1.6	4.5	7.6	1.7

synthetic action of sucrose and the store of carbohydrates in the grain go up with the increase in the oxidizing activity of ascorbinase.

From the data referred to it is obvious that in peas the complex of carbohydrates is not definitely related to earliness. The varieties of peas increase in carbohydrates in inverse proportion to their increase in proteins. The reason, which may be put forward to account for this fact, is that the products of photosynthesis are in greater amount used up when the synthesis of proteins is enhanced and this imposes limits upon the store of carbohydrates. And again, the increased synthesis of reserve carbohydrates sets up a deficiency of plastic materials in the plant limiting its store of proteins.

The results here reported entitle us to advance the assumption that the different share of enzymes in the carbohydrate and nitrogen metabolism of different varieties may be a factor in the different productivity of the latter with regard to storage of matter in their tissues. To clear up all these questions is the object of the work which is now being run in the laboratory.

Conclusions

1. The later the variety of pea, the higher is the synthetic activity of its protease and the greater is the ratio of protein in its grain.

2. The storage of protein in nitrogen-storing plants and the storage of sugar in plants of the carbohydrate type are governed by the same regularities: the high synthetic activity of the enzyme, acting as catalyzer in convertion of the fundamental groups of reserve matter, is responsible for the increase in storage of the respective groups and is related to the lateness of the variety concerned.

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BIOCHEMISTRY

ACTION OF HI ON CYCLOGLYCYL-GLYCIN AND THE BIOLOGICAL SIGNIFICANCE OF THE PRODUCT OBTAINED

By V. S. ISUPOV

(Communicated by L. A. Orbeli, Member of the Academy, 22. II. 1939)

Combinations of iodine and other haloids with the products of protein degradation and synthetic products of protein nature offer considerable theoretical and practical interest. There are many chemical methods and conditions of iodizing various amino-acids, peptides and cyclopeptides which have not yet been turned to account.

In the sphere of chemical and biological investigations a wealth of remarkable data on the significance of iodine-containing organic compounds, especially tyroxin and diiodtyrosin, has been obtained since the discoveries of Kendall and Harlington were made, but no works have been known dealing with the synthetic preparation of biologically active iodine-containing compounds which would be effected by acting on a cyclopeptide (diketopiperazine) with HI alone.

A comparative study of various reactions of iodine with cyclopeptides, peptides and aminoacids gave us the possibility to develop a simple method for iodizing the cycloglycyl-glycin with HI under mild conditions. The cycloglycyl-glycin was introduced for that purpose into tubes of molybdenum glass, half filled with 25% hydroiodic acid. This being done, the tubes were sealed up and placed into boiling water bath to heat them. The experiments have shown that the best conditions for the rection are obtained by heating the tubes during 6 hours and with a concentration of 50 cm^3 25% HI per gram of cycloglycyl-glycin. Under these conditions the whole of the initial material enters into reaction and is entirely iodized.

After cooling the tubes were opened. Thereby gaseous products were invariably observed to flow out. The liquid obtained upon heating the tubes had a colour of a concentrated alcohol solution of iodine. It was subjected to distillation in order to remove the excessive iodine, hydroiodic acid and water. The residue left upon distillation contained free iodine alongside of the iodized product. It was diluted with chloroform and the free iodine passed into the solution. Repeated distillation was made to remove the chloroform and its dissolved iodine. But the residue contained a small amount of free iodine still. One was compelled once more to add chloroform for dissolving the free iodine left and to proceed to the distillation again. Though repeatedly washed in chloroform and distilled, the residue would not give off its brownish colour entirely. Further on acetone was added to the residue; the flasc with its contents was heated with a reflux condenser until the boiling

point of acetone was reached, and the solution after boiling was put in a beaker for crystallization. The mixture was cooled in ice. Some minutes later aciculate crystals appeared. Isolated crystals were thrice recrystallized from the acetone, a white crystalline substance was obtained. The purified crystalline substance gave a distinctly positive reaction for iodine, when acted upon by a solution of chloramine.

The procedure described when repeated again and again proved to give recurring results. The iodine-containing crystalline product is readily soluble in cold water. The water colourless solution is perfectly clear and does not change in colour for a long time. It is worthy of note that the average yield of the iodized product is 50% of the initial material.

In order to estimate the ratio of iodine in the preparation obtained, the latter was analyzed in two ways: after the method of Gautier (1) and that of Leipert (2). The samples dried to constant weight and analyzed by the above methods showed enhanced figures of iodine. This made us to check the content of iodine by other two methods (Folgart and Stepanov). On determining its melting point, we observed the iodized product to disintegrate at 224°.

The analysis of the substance for its building up elements gave the following results:

Content of iodine

a) According to Folgart. To the samples dissolved in a small volume of water excessive silver nitrate was added, with a titre 0.02327 g/ml $AgNO_3$. The excess was titrated with rhodanic potassium. 1) 0.0973 g; 2.50 ml $AgNO_3$; 44.67% I. 2) 0.1345 g; 3.44 ml $AgNO_3$; 44.47% I.

b) According to Stepanov: 1) 0.2228 g; 5.65 ml $AgNO_3$; 44.10% I.

c) Estimation of carbon and hydrogen: 1) 0.1175 g; 0.0814 g CO_2; 0.0517 H_2O. Found: C—18.90%, 18.60%; H—4.92%, 4.64%.

d) Estimation of nitrogen according to Duma: 1) 0.1234 g; 12.2 ml N_2; H=748.7 mm; t=14°C. 2) 0.1026 g; 10.57 ml N_2; H=754.5 mm; t=17°C. Found: N—11.40%, 11.80%.

Average data: I—44.57%, C—18.75%, H—4.78%, N—11.60%, O—20.30%. Total 100%.

On ground of the data obtained by analysis the formula $C_9H_{21}N_5O_7I_2$ may safely be suggested for the compound in question.

We failed in estimating the molecular weight of the compound, either cryoscopically or according to Rast, on account of the dissociation of iodine.

The nature of the product analyzed would not be established thoroughly without disiodizing it in order to obtain the possibility of interpreting the reaction which had taken place. Purifying of the preparation and its analysis lead to a conclusion that the iodine failed to enter the ring of cycloglycyl-glycin or other compound derived, in response to the action of HI. Heating of the preparation in ethyl alcohol for one hour with a reflux condenser resulted in splitting off the iodine, and further repeated boiling removed it entirely from the preparation.

The disiodized product presented crystals of white colour, readily soluble in water and after their form looking like bundles. The following data have been obtained from its analysis: 1) 0.0809 g; CO_2—0.1004 g; H_2O—0.0595 g. 2) 0.0783 g; CO_2—0.0963 g; H_2O—0.0599 g. 3) 0.0810 g; N_2—12.3 ml; t=19°C; H=748.1.

Found: C—33.85%, 33.54%; H—8.23%, 8.56%; N—17.03%.

Average content: C—33.69%, H—8.39%, N—17.03%, O—40.89%. Total 100%.

The molecular weight of the disiodized product determined cryoscopically proved to lie near 175. Calculated from the analytical data the composition of the disiodized product is $C_4H_{10}N_2O_4$.

The formation of formaldehyde and methylamine has been detected in it by means of various quantitative reactions. The procedure was as follows. Excessive ammonia, potassium ferrocyanide and hydrochloric acid were added to the product dissolved in water. Crystals characteristic of hexamethylene-tetramine (rhombic and six-sided tablets distinctly visible under the microscope) resulted from this reaction, which fact pointed to the presence of formaldehyde. Another reaction for the presence of formaldehyde was run with a sulphate solution of diphenylamine. Aqueous solution of disiodized product gave a ring of emerald green in the test-tube. Similar positive reactions were obtained in control tests with a solution of pure formaldehyde.

Methylamine was detected by adding platinum chloride to a water solution of disiodized product. Yellow prismatic crystals precipitated. They were distinctly visible on the object glass of the microscope, pointing toward the presence of methylamine. Addition of picric acid to the disiodized product dissolved in water produced a yellow crystalline precipitate in the form of long laminae, thus furnishing an example of typical reaction for methylamine.

In order to interprete the structure of the disiodized product obtained through heating cycloglycyl-glycin with HI in a sealed-up tube and whose composition proved to be $C_9H_{21}N_5T_2$, this was, as we have mentioned before, disiodized and the result was a substance of the composition $C_4H_{10}N_2O_4$, the tetrahydroxypiperazine. The iodized product consists of two particles of tetrahydroxypiperazine bound mutually by the group HN—CH.

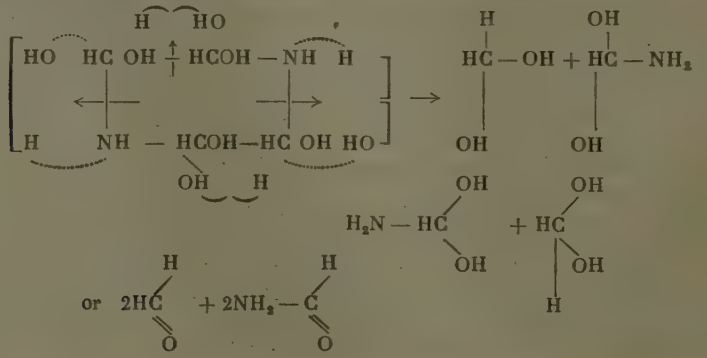
The descent of the disiodized product from cycloglycyl-glycin may be traced as follows:

I s t a g e. The conversion of cycloglycyl-glycin dihydrate into tetrahydroxypiperazine:

$$\begin{matrix} H_2C-C(OH)_2-NH \\ NH-C(OH)_2-CH_2 \end{matrix} \rightarrow \begin{matrix} HCOH-CH(OH)NH \\ NH-CH(OH)CHOH \end{matrix}$$

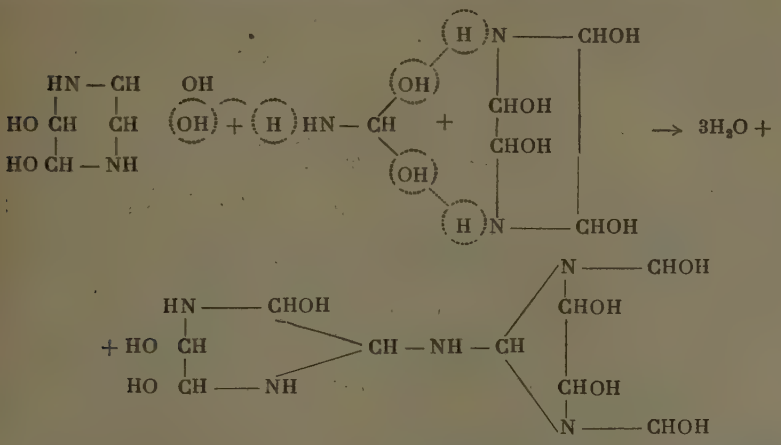
Dihydrate of cycloglycyl-glycin — Tetrahydroxypiperazine

II s t a g e. Formation of the link—NH—CH=. This link is a carbaminol (carbamine aldehyde) formed on the oxydation of the tetrahydroxypiperazine: the latter disintegrates into two particles of formaldehyde and two particles of carbamine aldehyde.

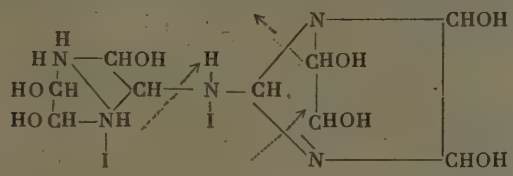


The formaldehyde is readily formed from the tetrahydroxypiperazine as it may be concluded from a number of reactions.

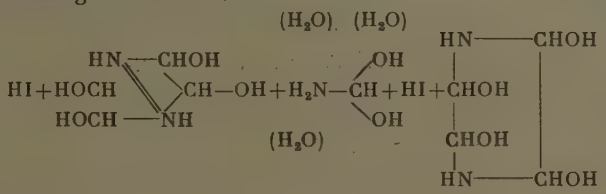
III stage. Condensation of two particles of tetrahydroxypiperazine with a particle of carbamine aldehyde when the latter is forming:



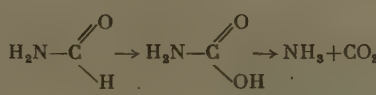
IV stage. Addition of two particles of HI. Since the product is rather easy to disiodize and combines not with three but only two particles of HI, it should be supposed to have the following structure:



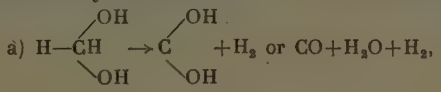
As to the process of disiodizing, it may be conceived as a hydrolytical splitting according to the reaction:



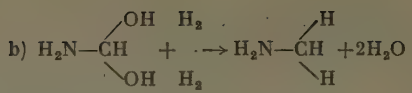
The carbamine aldehyde is either oxydized into carbamic acid and falls into CO_2 and NH_3:



or else it suffers a reduction by hydrogen, which is forming on the disintegration of the formaldehyde into carbon monoxide and hydrogen:



or into:



methylamine and water.

The factor in oxydizing is the hydrogen peroxide created in the following way:

$$HI+H\cdot OH \rightarrow HIO+H_2$$
$$HIO \rightarrow HI+O \text{ or } HIO+H_2O_2 \rightarrow HI+H_2O_2$$
$$2HI+H_2O_2 \rightarrow 2I+2H_2O.$$

Amino-acids and polypeptides have proved unable to form iodized derivatives with HI under the conditions described. The reaction according to V. Isupov and V. Sadikov, which we have referred to in the present paper, will not take place in the absence of peptin (dioxypiperazine) ring. Further experiments have to show whether similar products will or will not form with substituted peptins, such as cycloalanil-alanin, cycloleucin-valin and others.

It may be added here that the entire initial material, the cycloglycylglycin, was obtained from glycocolle when the latter mixed with glycerine in tubes of molybdenum glass was heated in an autoclave. Heating was run at 200°C for 2 and 2.5 hours.

At the end of our work experiments on animals were carried on. They were of preliminary nature indeed. It has been established that blood pressure is reduced in a cat by introduction in its veins of 0.03 of iodized preparation. The disiodized preparation, on the contrary, fails to produce any appreciable effect on the blood pressure of cat when injected in the same dose, and so it is also with the diiodtyrosin. Introduced into the cavity of axolotl in a dose of 0.004 the iodized product transforms them into amblystomes in 38 days. The biological action of the preparation will permit to prove their significance in medical practice.

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MINERALOGY

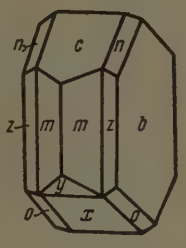
HYALOPHANE FROM THE SLYUDYANKA RIVER

By P. V. KALININ

(Communicated by A. E. Fersman, Member of the Academy, 2. III. 1939)

The phlogopite deposits of the Slyudyanka district in Southern Cis-Baikalia have been studied in detail and are well known in the mineralogical literature. There are no indications, however, as to the presence of feldspars in the phlogopite veins, though this group of minerals is rather extensively represented here and plays a considerable rôle in the genesis of deposits.

The feldspars of the phlogopite veins are rather manifold; just now they are being studied by the author and soon they will be described in detail. In the present communication the author would like to touch upon a variety containing a high per cent of BaO.



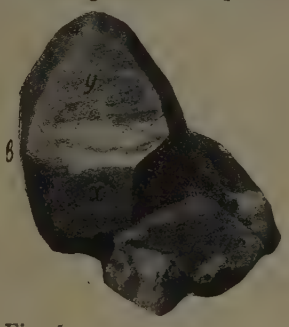


Fig. 1.

In 1935 when mapping the phlogopite veins at the Slyudyanka deposit, the author encountered several peculiar crystals of feldspar, distinguished from the ordinary mineral by their bluish-grey colour and a high transparency. A preliminary test showed the presence of barium, the analyses made showed a rather high BaO content (from 5.5 to 7%). The largest and beautifully developed crystals were found in the phlogopite calcite veins No 2 and 47 of mine No 1.

The phlogopite calcite veins of mine No 1 are lying among pyroxene amphibole gneiss, intersecting these almost across the strike. Schematically the structure of the veins is as follows: the central portion is filled with coarse-grained yellow or pink calcite, large (up to 1 m) perfectly developed crystals of blue apatite being enclosed in the latter. From the central calcite portion, phlogopite areas are symmetrically arranged, alternating with diopside-scapolite zones. In the places where no phlogopite is present, these zones are represented by large well preserved diopside (baikilite) and scapolite crystals, orientated in the direction of the vein, upon which crystals of hyalophane are sometimes developed.

Near the veins the enclosing rocks themselves have been subjected to a metasomatic replacement by diopside and scapolite and present a gradual transition to the rock from the vein body.

Thus, the barium feldspar is a later mineral compared to scapolite and diopside, but doubtlessly an earlier than phlogopite, apatite and calcite. The hyalophane crystals are somewhat elongated along the vertical axis.

The following faces are observed: (110), (130), (010), (001), (101), (111), (021) and (201) (figure). Dinstinct cleavage along (001). Colour—grey with a bluish tinge with pearly lustre. Hardness 6; specific gravity 2.701. Optical properties: N_g=1.540; N_m=1.538; N_p=1.534; $N_g N_p$=0.006; $2V$=—78°.

The chemical composition of hyalophane is given in Table 1.

Table 1

Oxides	Hyalophane from vein № 47, mine № 1. Analysis by Kalinin, 1934. (Moscow Geol. Prosp. Inst).		Hyalophane from vein № 2, mine № 1. Analysis by Langwagen, 1937 (Centr. Sci. Res. Geol. Prosp. Inst.)	
	Weight %	Molecular numbers	Weight %	Molecular numbers
SiO_2	59.85	0.996	59.54	0.992
TiO_2	traces	0.000	—	0.000
Al_2O_3	20.64	0.202	20.12	0.197
Fe_2O_3	—	0.000	0.17	0.001
FeO	—	0.000	0.29	0.004
MnO	—	0.000	—	0.000
MgO	0.21	0.005	0.04	0.001
CaO	0.86	0.015	0.20	0.004
BaO	6.92	0.046	5.75	0.038
K_2O	11.07	0.128	10.95	0.116
Na_2O	0.93	0.015	2.86	0.046
SO_3	not determ.	0.000	0.28	0.000
Loss on ignition	0.12	0.000	0.10	0.000
Hygroscopy	0.12	0.000	—	0.000
	100.82	—	100.30	—

Spectroscopic determinations give SrO about 0.5%. Recalculation to molecules of constituents is given in Table 2.

Table 2

Constituents	Formula	Analysis № 1, %	Analysis № 2, %
Orthoclase	$K_2Al_2Si_6O_{16}$	73.78	62.45
Albite	$Na_2Al_2Si_6O_{16}$	8.65	24.97
Celsian	$Ba(Sr)Al_2Si_2O_8$	13.25	12.98
Anortite	$CaAl_2Si_2O_8$	4.32	—

From Table 2 it may be seen that the investigated feldspar contains about 13% celsian molecule, which justifies it to be regarded as hyalophane, from its composition closely approaching the hyalophanes from the Manganese Mines of Yakobsberg in Sweden [1].

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MINERALOGY

HERRENGRUNDITE FROM THE USPENSKY MINE IN KAZAKHSTAN

By F. V. CHUKHROV and V. M. SENDEROVA

(Communicated by A. E. Fersman, Member of the Academy, 5. III. 1939)

Herrengrundite as a new mineral species was first described by Brezina in 1879 (2) and in 1880 Szabo (3) described the same mineral under the name of urvelgiite. The above authors studied specimens of herrengrundite from the old copper deposit Herrengrund, belonging at present to Czechoslovakia. Up to now this deposit has been the only herrengrundite deposit abroad. The first findings of herrengrundite in USSR are referred to 1938, when this mineral was detected by F. V. Chukhrov in the Uspensky copper mine, which is located near the station Nelda of the Karaganda—Balkhash railroad-line.

Segregations of herrengrundite in the Uspensky mine have been detected at a depth of 84 m from the surface in a cut of ort No 1. The mineral is a typical new formation, precipitated from descending waters in old abandoned workings both upon their walls and on the floor, and on the remains of timbering. The most abundant quantities of herrengrundite are observed in places, where the waters, from which it is deposited, are percolating most slowly.

Herrengrundite usually segregates in the form of films and crusts, the thickness of the latter attaining 5 mm. The crusts consist of thin foliae orientated perpendicular to the surface upon which they are arranged. The surface of the crusts is occasionally smooth and partly rough with showings of thin ribs; in part these crusts have a lamellar structure. The colouring of herrengrundite is bright bluish-green at the surface and at the breaks of the aggregates it is light apple-green. It is difficult to determine the hardness of individual foliae owing to their minute size; the hardness of the aggregates is about 1. In acids easily decomposed. When dissolved in ammoniac, a residue is precipitated belonging to calcium sulphate. When ignited in a closed tube, yields water.

An analysis of herrengrundite from the Uspensky mine made by V. M. Senderova gave the results shown in the following table. A small loss in the total may be explained by the presence in the mineral of lead, zinc and manganese, which have been spectroscopically detected, but failed to be determined quantitatively owing to scarcity of material.

The formula of the mineral analyzed is $CaO \cdot 4CuO \cdot 2SO_3 6H_2O$, which corresponds to the established formula of herrengrundite. The X-ray diagram of the herrengrundite from the Uspensky mine was found to be identical to that of the herrengrundite from Herrengrund.

Constituents	Analytical data in %	Molecular quantities	Ratio of molecular quantities
CuO	48.42	0.608	4.00
CaO	9.14	0.164	1.08
MgO	traces	—	—
SO_3	24.28	0.303	2.00
H_2O^+	16.52	0.918	6.04
H_2O^-	traces	—	—
$Al_2O_3+Fe_2O_3$	0.34	—	—
SiO_2	0.62	—	—
	99.32		

For the genetic characteristic of herrengrundite its mode of occurrence in Herrengrund is of interest. According to J. Szabo, malachite, azurite, gypsum and pitticite have been found in paragenesis with herrengrundite. Occasionally in a close proximity with herrengrundite tetrahedrite has been observed which, together with chalcopyrite, is regarded as a source of copper, fixed in secondary minerals. The ore veins of Herrengrund lie under micaceous schists and are overlain by grey-wacke. Over the latter Triassic limestones are lying which, probably, yielded the calcium which served for the formation of herrengrundite and gypsum.

The formation of herrengrundite in the Uspensky mine is fully connected with the chemistry of the waters of inhibition which, according to Yagovkin ([1]), are enriched here with calcium bicarbonate and sulphate. The problem as to the age of the herrengrundite described is of interest. As it has been shown above, the specimens studied by the authors are taken from a depth of 84 m. According to Yagovkin, the workings at the Uspensky mine were begun in 1855. The horizons from 63 to 108 m were worked during the period 1899—1904. It is very probable that just at that time the working was made, from which originated the herrengrundite described by the authors. In connexion with this it may be shown that the age of the oldest segregations of herrengrundite in this working is not more than 40 years.

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MINERALOGY

ON THE PROBLEM OF THE CHEMICAL COMPOSITION OF ZIRCONS

By E. KOSTYLEVA

(Communicated by A. E. Fersman, Member of the Academy, 10. II. 1939)

A new analysis of zircon from the south Urals, with a total of 104.53 containing about 2.56% Na_2O and 69% ZrO_2 [according to prof. Jakob's proposition (1) Zr is in the lowest degree of oxidation], made the author reexamine critically the old and the new analyses of the zircons in the USSR, in the course of his studying the geochemistry of zircons, and perform special alkali determinations for some, and full analyses for three specimens. These analyses, which were performed by Pereverzeva, neither revealed considerable quantities of alkali in these samples nor found the total of analysis to exceed 100, and gave therefore no ground for supposing Zr to present in the form of Zr_2O_3 and ascribing a somewhat different formula to these zircons. The analyses were performed for zircons from the following layers, and having the following physical properties (Table 1).

Table 1

Layer	Colour	Specif. gravity		Powdergram
		raw	calcined	
South Urals	Usually brown	4.45	4.79	Usual one for zircons
Khibin tundra . . .	Light brown, lightly yellow	4.14	4.23	Do
Ceylon	Green transparent	3.96	3.90	After calcining ZrO_2 and SiO_2 lines

When the analyses were finished, there arose a number of new problems which should be looked into: 1) the problem of the isomorphous substitution of Zr by elements having valencies other than that of Zr, in connexion with the formula of the zircon; 2) that of the various solubilities of zircons in hydrofluoric acid; 3) that of the U content, as being the cause of the metamict decomposition of zircon.

1. The Presence of Elements isomorphously substituting Zr, Si and may be O in zircons should obey the law of compensational isomorphism. From this point of view if the formula $ZrSiO_4$ is correct, the presence of elements of the same valency, as Hf, Th, U (probably

U····), in the zircon lattice does not need electrostatic balance compensation and does not provoke conjugate isomorphous substitution of the anion; the presence of elements of other valency, such as TR, Fe, Ca, Nb, which are often found in zircons, should respectively make necessary the presence of an anion of other valency. Of the anions which displace Si in zircons, analysis shows the presence of Al (Be), Fe, and lately P([2]). With the exception of Fe these admixtures are usually so insignificant in zircons that they are not taken into account in calculations, but in the varieties of zircon—malacon and zirtolite—both the isomorphous admixtures' content and the discrepancy in the ratios of ZrO_2 and SiO_2 are quite great, and have not been expressed in formula up till now. Analysis No 2 from the Khibin tundras, which was performed by Pereverzeva, is of great interest in this connexion, since it contains 1.48% CuO and no other anions except Si (neither P nor Al was detected). Special tests for sulphur on two freshly chosen samples of zircon No 2 from the same layer, performed by L. B. Tumilovich, showed the presence of S both times; in one case a quantitative test gave 0.75% SO_4; evidently, in this case the presence of S is connected with the complex (SO_4) and not with the anion S'', and we can assume that $CaSO_4$ molecules are present in zircon isomorphously, or that we have an indefinite solid solution of $CaSO_4$ in $ZrSiO_4$ as solvent. The fact that S was found for the first time by us in zircon, on the basis of the theoretic premises given above, shows that analysis of minerals of the zircon group have not always been complete and accurate, and deviations from the ratio $ZrO_2 : SiO_2 = 1 : 1$, given by the theoretic formula, are most probably due to the incompleteness and incorrectness of the analytical data for the minerals of this group. According to the data of the roentgenochemical laboratory, a roentgenochemical analysis of zircon No 2 from Khibin tundra shows the presence of Sr, Y and Nb besides quadrivalent Hf and Th; Y and Nb are present in insignificiant amounts. The main difficult points in analyzing minerals of the zircon group are: the difficulty of decomposing the mineral, the methodic difficulty of the direct determination of small quantities of Al in the presence of great quantities of Zr, the estimation of Nb and Ta and other elements, such as F, S.

2. The Varying Solubility of Zircons in Hydrofluoric Acid is mainly due to the metamictic decomposition of zircon. This is confirmed by the fact that the green zircon from Ceylon decomposes (analysis No 3) practically completely in HF, sp. gr. 3.96, with a powdergram of ZrO_2 and SiO_2 after calcining. But metamictic disintegration seems to be not the only cause of decomposability of zircons in hydrofluoric acid, and not all «normal» zircons resist the action of HF equally. The insolubility of Ural zircon in HF, as established in publications ([2]), was confirmed by us also in the case of zircon from mine No 26 of the south Urals. Considerable solubility (75%) of zircon No 2 from the Khibin tundras, with a normal zircon powdergram, was unexpected. Experiments on treating macroscopically homogeneous zircon crystals from this layer by evaporating them for 1 hour with HF in a bath showed that the crystal structure becomes heterogeneous, as a result of the various actions of the acid. After treating with acid, irregular sections appear in the centre of the crystal, which lose their transparency, receive a mother-of-pearl bluish colour, and become susceptible to the action of acid; these sections are arranged in zones towards the periphery, and alternate with the transparent unchanged zircon layers, which form the outermost layer of the crystals. In thin sections the heterogeneity in the structure of the crystals becomes sometimes apparent in the portions with slightly changed structure and in zones with other more interferentional colour. A number of zircon crystals from peg-

matite veins of the south Urals, which are semi-transparent, but macroscopically homogeneous, also react differently with HF and give zonal crystal structure after being treated. Tiny transparent hyacinth red zircon crystals from the south Urals did not show this phenomenon. Heterogeneity in the crystal structure of zircon is proved by luminescence analysis also—by subjecting them to the action of ultra-violet light rays. Zircons luminesce heterogeneously; dark brown and velvet-black sections alternate with bright yellow ones. After treating the zircons mentioned above with hydrofluoric acid, the predominating colour is a brownish yellow-ochre. Whole zircon crystals from the Khibin tundras, velvet-black under the ligth of ultra-violet rays, become a brigth yellow-ochre after the treatment.

The facts mentioned above show that zircons do not resist HF equally and that this phenomenon is connected with the heterogeneity in the composition and lattice structure, which was observed by us for some zircons.

3. Cause of the Metamictic State of Zircons and of the Decomposition of Their Lattices. The decomposition of zircon lattices into ZrO_2 and SiO_2, as has been established in the publications of the last few years ([3-6]), makes the specific gravity of the zircons decrease to 3.9 and brings anomaly and changes into the physical properties (hardness, brittleness, solubility). The problem of the cause for such a complete decomposition has not yet been solved finally n literature. In this relation the analysis of zircon No 3 from Ceylon, which possesses the typical properties of metamictic zircon, is of great interest, since it shows the presence of 0.65% U_3O_8. Determinations of U in three different zircons from Ceylon (by means of α-rays, and roughly ground in comparison with the pulverized standard, 0.1% U_3O_8) confirm the U-content in metamictic zircons. The determinations were performed by V. I. Baranov (Table 2).

Table 2

Layer	Colour	Equivalent U_3O_8 in %
Ceylon zircon	Light . . .	0.21
» »	Green . . .	0.61
» »	Black . . .	1.19

Thus, there is no doubt that metamictic decomposition is brought about by α-radiation of U which is isomorphously enclosed in zircon. Ceylon zircons, containing small quantities of U (0.65—1.14), are decomposed into ZrO_2 and SiO_2; the degree of this decomposition is determined by the quantitative U-content; in this case α-radiation seems not only to make the lattice decompose into ZrO_2 and SiO_2, but also to cause the green and black colour, which is abnormal for the mineral. According to F. Lietz' papers ([7]), the phenomena of colour and colour variation in zircons are caused, just as in the case of the alkali halogenids, by the liberation of metallic atoms in the intersections of the lattice (F and F′ centres); it may be assumed that the latter is made easier by the influence of α-radiation.

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CRYSTALLOCHEMISTRY

CLASSIFICATION OF CLOSEST AND CLOSE PACKINGS

By N. V. BELOV

(Communicated by A. E. Fersman, Member of the Academy, 26. II. 1939)

Pauling (1) has been the first to show that the number of different closest spherical packings is not restricted to two widely known cases—the closest cubic and closest hexagonal ones, but their number is infinitely great. If closest packings are being composed of planar closely packed layers, vertically stacked on each other in such a way that the spheres of each new layer fit in the hollows of the underlying layer, then the alternative for the third layer to lead either to a cubic */ABC/A...*, or hexagonal case *AB/A...* is preserved for each of the subsequent layers, and this makes the number of closest packings infinite. The third simplest case, first shown by Pauling, is expressed by the formula */ABAC/ABAC/...* Here the period consists of four layers instead of two or three, as it is the case in hexagonal and cubic packings. According to Pauling, this new case of closest packing is realized in the structures of topaz and brookite. Later on closest topaz packing has been detected in swedenborgite, carborundum III, in the $MgZn_2$ phase.

Other cases of closest packing, encountered in natural and synthetic compounds: */ABCACB/ABCACB/...* (a period of 6 layers)—carborundum II, $K_3W_2Cl_9$, LaF_3 (natural tysonite, after which this packing may be called tysonitic); */ABACACBCB/A...* (9 layers)—aluminium carbide Al_4C_3.Still more complicated are the packings od carborundum I (15 layers) and $CaSi_2$ (18 layers).

Any infinite sequence (not necessarily periodical) of three letters *A*, *B*, *C*, in which two identical letters will nowhere stand side by side, determines a closest spherical packing with the same grade of space utilization (74.05%), as in the two principal cases. All such packings are characterized by a system of threefold axes, passing through each *A*, *B*, *C* position, i. e. in principle every closest packing is trigonal with an increase of symmetry in individual cases up to a sixfold (on account of screw axes—hexagonal, topaz, tysonitic) and cubic symmetry (on account of additional non-vertical threefold axes).

Beside the three-letter formulae (*A*, *B*, *C*), there exists another method to denote the closest packings, using only two letters (2). Each of the letters *A*, *B*, *C* is replaced by one of the letters *p* or *a*; if in the original formula the letter to be replaced is standing between two different letters, we use *p*, and if it is standing between two identical letters, we replace it by *a*. Thus, the cubic packing is represented by the formula *ppp...*, and the hexagonal by *aaa...* Therefore, the author prefers to use the letters *c* and *h*, in

which case cubic packing is expressed by *ccc*...., the hexagonal by *hhh*..., the topaz by *hchchc*..., the tysonitic by *hcchcchcc*..., $MgNi_2$—*hccchccchccc*..., Al_4C_3—*hhchhchhc*.., etc. Formulae of this type are especially convenient for the construction of lattice models, e. g., by the method of Pauling's polyhedra, inasmuch as it reflects (as we shall see) the characteristical features of every packing.

The deeper sense of the second method lies in the fact that in all closest packings the composing spheres, from the character of their nearest surroundings (e. g. from the arrangement around each sphere of 6 octahedral

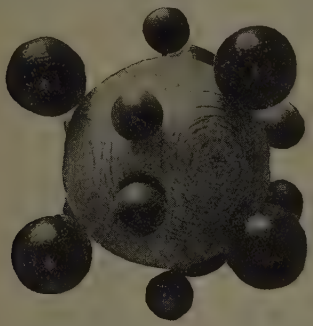

Fig. 1. Spheres of a closest hexagonal packing.

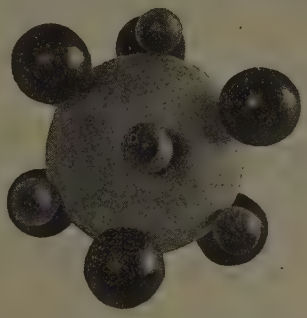

Fig. 2. Spheres of a closest cubic packing.

or 8 tetrahedral «holes»), may be of two and only two types: either of the *h*-type (with surroundings of a hexagonal character—Fig. 1) or of the *c*-type (with surroundings of a cubic character—Fig. 2). In the hexagonal packing all the spheres are of the *h*-type, in the cubic of the *c*-type, in other packings layers of spheres of one type alternate with layers of spheres of another type, according to the formula of the letters *c* and *h*.

The method in general use recurring to superposition of planar closest layers is not quite convenient, since it does not permit to represent the packings by a planar drawing, but requires a perspective representation.

Fig. 3. Simple trigonal packing.

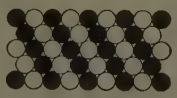

Fig. 4. Closest cubic packing (*ccc*...).

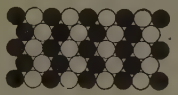

Fig. 5. Hexagonal closest packing (*hhh*...).

There is, however, another well known method to obtain the cubic closest packing. Proceeding from an initial planar not closest packing, namely a square packing, we superpose the layers in such a way that the spheres of each new layer fit in the hollows of the preceding layer. It is a simple matter to show that in this case too closest packing is derived from a planar not closest one, owing to a deeper sinking of the spheres of one layer into the other.

The purpose of the present note is to show that this method, employed up to now exclusively for cubic packing, may be used also for all the cases of closest packings.

Suppose the square layers in a cubic packing built up by the above method to be arranged vertically, and the following interpretation becomes possible. Let us proceed from initial close-packed layers superposed (in a direction perpendicular to the drawing plane) vertically (the spheres of

every new layer upon the spheres of the underlying layer—the so-called simple trigonal packing—Fig. 3); divide this packing into vertical layers of one sphere thickness; shift these layers relatively to each other for $^1/_2$ translation $= {}^1/_2$diameter of the sphere (Fig. 4—black spheres at the zero level, white spheres at 1), and finally, compress the packing to such an extent that the spheres of the displaced vertical layers would fit in the hollows of the neighbouring vertical layers.

It will be easy to demonstrate that the ultimate result is a closest packing (74.05% of the whole space being filled with spheres) even in the case, when the initial simple trigonal packing is divided into vertical layers (thickness 1 sphere) not along straight, but along undulating, zigzag lines.

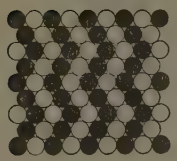

Fig. 6. Topaz packing (*hchchc*...).

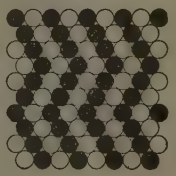

Fig. 7. Tysonite packing (*hcchcc*...).

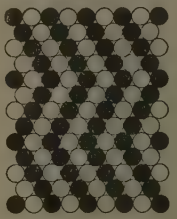

Fig. 8. $MgNi_2$ packing (*hccchccc*...).

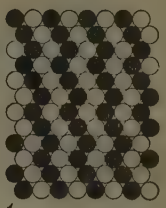

Fig. 9. Al_4C_3 packing.

Fig. 5 shows the scheme of formation of a closest hexagonal packing. In the mutually displaced layers the number of zigzags is maximum. Each sphere is the vertex of a zigzag. Length of the zigzag—2 spheres (more exactly, 1 sphere diameter). Formula *hhh*...

In the case of a topaz packing (Fig. 6) the zigzags are symmetrical as before; length of the zigzag—3 spheres (2 diameters). The spheres forming the

Fig. 10. Closest cubic packing(final scheme).

Fig. 11. Anion packing of valentinite.

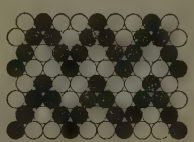

Fig. 12. Anion packing of antimonite.

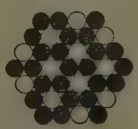

Fig. 13. Anion packing of andalusite.

apexes of the zigzags alternate with the spheres arranged in the middle of the zigzag lines. The number of spheres of both types is the same. Formula *hchchchc*...

In tysonite packing (Fig. 7) the length of the zigzag is 4 spheres (3 diameters). The zigzags are symmetrical, but the number of the spheres at the apexes of the zigzags is half as great as the number of spheres in the middle portions of the zigzag lines, according to the formula *hcchcch*...

In the $MgNi_2$-packing (Fig. 8) the length of the zigzag is 5 spheres (4 diameters). The number of spheres at the apexes is a third of the number of spheres in the middle portions of the lines, according to the formula *hccchccch*...

In this way we arrive at one more method of interpreting the letters *h* and *c* in the formulae: the sphere *h* determines the presence of a zigzag at the given point and is arranged at the apex of that zigzag; the sphere *c* indicates the absence of a break at the given point of that line.

In the Al_4C_3 structure (Fig. 9), according to the formula *hhchhc*..., the zigzags cease to be symmetrical; one side of the zigzag is composed of 2 spheres, the other of 3. In carborundum I (*chchcchchcchchc*...) the links of the zigzags are composed in turn of 3 and 4 spheres.

It may be easily seen that actually for all cases of closest packing this method leads to a clear planar picture of the packing, according to the formula of the letters *h* and *c*. It is characteristic of our method that we proceed from planar layers (trigonal packing) with threefold axes perpendicular to the drawing plane and arrive at a packing with threefold axes in the drawing plane (parallel to the direction of the chain axes).

This simple scheme of all the closest packings becomes somewhat complicated from the following circumstance. All the schemes just examined show only the initial division of a simple trigonal packing into vertical layers. But after the division has been accomplished and the layers are displaced for $^1/_2$ translation, they are still pushed into each other, and this results in a disturbance of the projection symmetry perpendicular to the former threefold axis. Equilateral triangles with spheres lying at their apexes cease to be such. If in the initial system the sphere of layer *2* (Fig. 3)

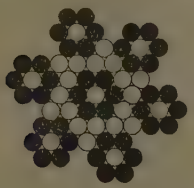

Fig. 14. Anion packing of apatite.

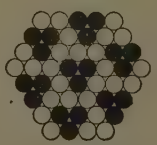

Fig. 15. Anion packing of benitoite.

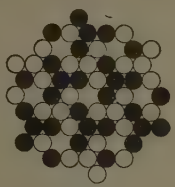

Fig. 16. Anion packing of phenacite.

be shifted to the right (or to the left) from the spheres of layer *1* for $^1/_2$ diameter, then, as it is easy to calculate, the displacement becomes greater (in the direction of the zigzag) up to $^2/_3$ diameter as a result of compression, i. e. the angles of the zigzag sides with planes of the closest packing cease to be arccos $^1/_2$=60°, but are arccos $^2/_3$=∼49°.

In cases when, as in most schemes examined, the zigzags are symmetrical (individual letters *h* divide equal numbers of letters *c*), an additional shift to one side will be compensated immediately by an additional shift to the opposite side, and thus the initial schemes remain ultimate in these cases.

If, however, the zigzags are not uniform, this circumstance, without changing the picture qualitatively, changes it with regard to the period of recurrence. Thus, in the case of cubic packings (*ccc*...) the additional displacement produced by each letter will result in the period of recurrence, being not (2 : $^1/_2$)=4 layers, as it would follow from the initial scheme (Fig. 4), but (2 : $^2/_3$)=3 layers, as it is given by the ultimate scheme (Fig. 10). In the case of Al_4C_3 the letters *hh* mutually cancel their effects, and the whole shift is to be ascribed to the letters *c* acting in one direction. The result is that we shall come for Al_4C_3 to a period of 9 layers instead of 12.

The reduction of all closest packings to a simple trigonal packing by arranging the spheres of the latter at two levels has also a general meaning, because most important cases of packing, which cannot be reduced to the closest, can be obtained from the trigonal packing by the same method. Figs. 11—16 show the schemes of anion packings of valentinite (Sb_2O_3), antimonite (Sb_2S_3), andalusite (Al_2SiO_5), apatite ($Ca_5P_3O_{12}F$), benitoite ($BaTiSi_3O_9$) and phenacite (Be_2SiO_4).

It will be obvious that with a reservation of the main principle of arrangement of the spheres at two levels, neither of these packings may be the closest, since in the pattern of each a closed figure composed of spheres is observed at one and the same level (traingles in Sb_2S_3, hexagons in andalusite and both in apatite).

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HISTOLOGY

ON THE IN VITRO CULTIVATION OF PERIPHERAL NERVE PARTICLES

By N. G. KHLOPIN

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The study of the tissue elements of peripheral nerves, and in the first place of Schwann's cells, by the method of tissue cultures, is a necessary link of the chain formed by the cycle of studies on the auxiliary components of the nervous system made by my collaborators and guided by me ([8],[1],[14]).

A summary exposition of the general trend of these researches and of some general conclusions drawn from them has been recently published in this periodical ([8]).

The auxiliary components of the peripheral nervous system have scarcely been studied up to the present by means of the tissue culture method. Besides the study of Ingebrigtsen, published over 20 years ago ([2]), which concerned the explanation of particles of peripheral nerves in the state of Waller's degeneration, the literature on the question contains but quite scarce data regarding the growth of embryonic lemmoblasts ([3],[7]) and observations on the degenerative processes in isolated particles of peripheral nerves ([5],[6]).

The purpose of the present work was the study of the general character of growth of Schwann's cells and their isolation in pure culture. As material for the study we took n. vagus particles of new-born and young rabbits 1—20 days of age, as well as of rabbit embryos a short time before birth and of pig embryos. Moreover, for some of the experiments rabbits aged 1—3 months were subjected 7—10 days before the explantation to a double ligature of one of the n. n. vagi. In most cases the particles of n. vagus before explanting them were liberated as far as possible of their connective tissue envelopes.

The tissue elements of n. vagus in the explants are distinguished by an intense growth and proliferation and may be cultivated with success for a prolonged period of time outside the organism. At the early stages of cultivation besides a more or less intense growth of the connective tissue elements, one observes also an active proliferation and growth of Schwann's cells. Proliferating Schwann's cells in many cases suppress the growth of connective tissue and after some transfers can be isolated in pure culture. Schwann's cells possess a marked fibrinolytic capacity and may be distinguished on the ground of the total of their transformations at different stages of cultivation, followed by us for over two months, from every other tissue element studied up to the present, including the derivatives of the mesenchyma. The histological structures of the zone of growth, which are most characteristic for Schwann's cells and are observed at every stage of cultivation, are

combinations of elongated cells—in some cases lying on one plane, in others more massive and rather complicated—which at the periphery of the zone of growth pass into loosely lying tracts or chains of cells drawn into the length, often filament-like with bead-like thickenings in the region of the nucleus. The tracts of closely linked Schwann's cells may grow in bunches, or break into isolated elements of varying, frequently also filamentous shape. Some filamentous cells may form long fine cytoplasmatic side outgrowths, which not infrequently become linked with the neighbouring cellular tracts and form peculiar reticular structures. On relatively rare occasions these tracts of Schwann's cells are composed by broader elements. The Schwann cells are characterized by their ability to become particularly closely and continuously linked one to the other by their ends. Their lateral bonds are usually much less marked and are easily destroyed. In the cytoplasm of the dense complexes of Schwann's cells in preparations, stained by Heidenhain's iron haematoxylin, fine filamentous structures may be detected. Besides this, the combinations of Schwann's cells are able to form erythrophile filaments of a particular intercellular substance.

In the course of their transformations under the conditions of tissue cultures Schwann's cells may be encountered at the different stages of cultivation also under less characteristic shapes of their existence, forming sometimes loose «mesenchyma-like», sometimes complex «epithelium-like» cellular combinations. Isoplated Schwann's cells may on such occasions acquire the state which is under certain conditions characteristic of their existence outside the organism—of tissue elements of the most different origin, i. e. the «fibroblast-like» shape. At earlier stages of cultivation part of Schwann's cells transform into amoeboid or spheric «microphage-like» elements. In the course of time the number of such «microphage-like» cells diminishes, as a rule at the later stages they are totally absent. The tissue nature and the specifity of the less characteristic existence forms of Schwann's cells may be—as in the case of many other tissue elements—correctly understood, only if the way of their formation and their further transformations are systematically followed throughout.

Combinations of mutually bound Schwann's cells show a doubtless likeness to the growing outside of the organism neuroglia of the optic tract, which we have studied somewhat earlier ([8]). Schwann's cells differ from the latter under the conditions of explantation in that the peripheral areas of the zone of growth formed by them are broken into peculiar narrow cellular tracts, often even thread-like, and into isolated cells. This less compact peripheral area of the zone of growth may cover a quite considerable surface, as compared with its dense-cellular central areas. This points but to the poorly developed ability of Schwann's cells to form complex structures. The particular characteristics of the transformations of the neuroglia of the central nervous system outside of the organism, which indicate, though in a peculiar and changed form, its kinschip to the epidermal tissue, are practically absent in Schwann's cells.

By the totality of their transformations outside of the organism Schwann's cells are histologically strictly determinated elements of peripheral neuroglia, incapable of progressive transformation into tissue elements of different origin. It seems the most correct to consider the growth forms of central and peripheral neuroglia as different modifications of a common neuroglial type of growth. Both the central and the peripheral neuroglia are genetically one tissue type, the different sections of which have lost in the course of the philogenetic process their original barrier or delimiting properties to an unequal degree and have acquired the properties of a specific tissue of the internal medium of the organism.

The histotypic characteristics of Schwann's cells revealed by the method of explantation allow to understand correctly the structures which characterize the so-called neurinomes, the nature of which remains discutable up to the present. The histoblastic potencies and histologic determination of Schwann's cells speak for the arising of the neurinomes out of peripheral neuroglial elements and against viewing them as perineural fibroblasts.

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HISTOLOGY

TRANSFORMATIONS OF POLYSTICHOUS EPITHELIUM OF THE TRACHEA IN REPARATIVE REGENERATION

By N. I. GOLSTEIN

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According to the natural system of the so-called epithelial tissues (Khlopin), the polystichous prismatic flittering epithelium of the respiratory ways belongs to the group of tissues of the epidermal type. This group of tissues is characterized in the vertebrates chiefly: 1) by its ability to produce polystichous structures and 2) by a more or less marked vertical anisomorphy in the condition of the differentiated tegumental layers. It has already been pointed out since long that in the process of reparative regeneration the polystichous epithelium of the respiratory ways passes through the stage of polystichous flat epithelium. Recently this has been confirmed by Strelin's tissue culture experiments and by the experimental studies of Garshin and his co-workers.

To check the ability of the epithelium of the respiratory ways to produce polystichous structures, as well as to prove the cambial significance of its basal layer (which as yet lacked sufficient clearness), experiments were arranged on rabbits, which were subjected to an open burn on a limited area of the tracheal wall. Observations on the dynamics of the transformations of the tracheal epithelial lining in the course of the process of reparative regeneration lasted from 6 hours to 60 days, thus permitting to establish a series of regularities.

We succeded to establish that in the pre-existing epithelium, in the immediate vicinity of the defect, there arises, particularly at the early stages, a gradual simplification (the so-called dedifferentiation) of the epithelial layer, which looses its polystichous structure, a gradual desquamation and separation of the specialized cellular elements, and the formation at the border of the defect of polystichous areas of varying thickness, composed by indifferent cells. A typical polystichous structure, as a transitory stage, was also found in the young regenerating tissue, covering the defect by the 5th day or somewhat later and producing subsequently a polystichous structure with a flittering tegument. Thus, the views of Khlopin as to the kinship of the lining of the respiratory ways with the epidermal tegument of the skin and of the anterior intestine are once more corroborated, as well as the opinion repeatedly expressed by Garshin, who emphasizes the necessity of distinguishing between the notion of genuine metaplasy of the respiratory epithelium and its so-called ateleplasy.

The character of the mitotic activity, which reached its utmost intensity on the 2nd and 3rd day after the instigation of the burn, was followed in the greatest detail. Both in the intact epithelium (as a very rare occasion, it is true) and at the earliest stage of the mobilization of the epithelium after its traumatization, individual division figures were observed only

in a few cells of the basal layer, at a considerable distance from the defect. During the period of the highest mitotic activity, a complete parallelism is established between the degree of rearrangement of the epithelium and the changes in the localization of the mitoses. Hand in hand with the progressing simplification of the epithelial layer, there is a penetration of the mitotic figures from the inferior cell layer to the upper ones, while in the areas lying nearest to the defect and composed by indifferentiated cells the mitoses in all the layers are distributed at random. The shift of the mitotic activity proceeds not only in the vertical direction, but along the layer as well, since alongside of the increase in the number of mitoses in the peripheral sections in their basal cells, a particularly high number of mitotic division figures is observed in the immediate vicinity of the defect. A similar displacement of the mitotic activity—though in the opposite direction—is noted in the course of the process of differentiation towards a complete restoration of the histological structure of the epithelial layer, when the function of proliferation is finally preserved only by the cells of the basal range. This observation confirms the presence of vertical anisomorphy, i. e. the cambial significance of the basal cells of the differentiated polystichous prismatic flittering epithelium.

The totality of the phenomena observed—the morphological characteristics of the marginal (bordering on the defect) zone of the epithelial stratum and the character of the mitotic activity in this zone—permit of forming an idea of the mechanism of the recovering of the wound's surface by the regenerating epithelium. The process of reparative regeneration of the polystichous tracheal epithelium is built up by two factors: the mitotic activity of those cells of the epithelial lining which are able to divide, mainly of the cells lying at some distance from the margin of the defect, and the active progression of the stratum as a whole in the region of the defect itself. The processes taking place in the epithelium are correlated with the underlying connective tissue, which provides for the growth and solidity of the young epithelial cover.

In the case of delayed regenerative processes there is a disorderly «inflammatory» growth of the epithelium with an overgrowing of the necrotized masses deep within the defect and the formation of cyst-like cavities (the manifestation of the eliminative function of the epithelium according to Garshin).

Usually within 5—7 days after the burn, the integrity of the epithelial stratum is already restored, while by the 15th day takes place also the restoration of the continuousness of the flittering epithelial cover.

The course of the reparative regeneration of polystichous tracheal epithelium depends on a number of secondary factors and does not fit within universal scheme. Yet, in spite of the variety of deviations, certain characteristic phenomena which reoccur in regeneration not only under the conditions of a whole living organism, but also in in vitro tissue cultures, point to the fact that they are hereditarily fixed specific properties of the given tissue. These properties permit of forming a judgment to the effect of a genetic relationship existing between the polystichous epithelium and the other epithelial structures of the epidermal type.

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HISTOLOGY

GROWTH AND TRANSFORMATION OF FOLLICULAR AND TEGUMENTAL CELLS OF THE OVARY OUTSIDE THE ORGANISM

By N. A. KOLESNIKOVA

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The purpose of the present work is to study the specific peculiarities of the growth of follicular and tegumental cells of the ovary outside the organism. Literary data on the cultivation of ovary tissue can be subdivided into two categories. One part of the authors ([4],[11]) believe that follicular cells in tissue cultures end by acquiring the features of connective tissue elements. The opinion of the other part of authors([1],[2],[6],[9],[3]) is that the follicular and lining cells of the ovary show the epithelial type of growth.

In the present work as material for investigation we used ovaries of rabbits, of one to 26 days of age, and ovaries of adult females. The cultivation was done according to the hanging drop method. A total of 30 series of experiments was carried out, the number of all cultures reaching about 500. Half of these were fixed and stained. The longest period of cultivation equalized 26 days.

Alterations in the explanted bit are observed already during the first 24 hours. The bit shrinks, dying-off cells separate from it and an active migration of amoeboid elements of mesenchymatous origin into the surrounding medium begins. In the bit, alongside of the usual degenerative alterations, there takes place a rise in the mitotic activity of cells of every kind. In the sexual cells mytotic divisions end, as a rule, after two days. The sexual cells are the first to suffer from degenerative processes. The chromatin in them gathers into granules and clots, connected by filaments with one another. At first the intranuclear content disappears, the nuclei seem empty; later ensues the decomposition of the cytoplasm. The follicular cells in the centre of the explant may keep viable for a relatively long time and divide by mitosis. 10—12 hours after planting one can see separate cells of connective tissue, which is mobilized before all others, to grow into the fibrin. Later on the growth of connective tissue may be stifled by that of other elements of the ovary. During the first 24 hours in the areas of the fibrin clot, bordering directly on the lining epithelium of the ovary, if the explant contains such epithelium, there appear small cavities due to liquefaction. The process of the liquefying of fibrin goes on at a relatively slow rate and is much less marked than in cultures, containing epithelial tissues of the dermal, entetic or nephrous type. The lining cells, beginning to grow out. move along the border of the cavity of liquefaction producing arch-like structures. Before

growing beyond the borders of the bit they may for some time grow over its surface. Yet, they never cover up considerable areas of the explant's surface, nor is the picture of its complete «epithelization» ever observed. If the explant is taken from the ovary of a new-born rabbit, in which case it is impossible to distinguish lining epithelial cells from follicular cell even on the original material, then by the end of the second or third, and particularly of the fourth day, alongside of the lining cells begin to grow follicular cells of the sexual tracts also, though it is not possible to draw a strict line between the lining and the follicular cells. The lining epithelium of the ovaries of elder rabbits, penetrating into the fibrin in the course of its extensive growth, also suffers alterations similar to those of the follicular cells and forms structures identical with those of the latter. In case follicles are growing at the periphery of the explanted bit, it is possible to see the follicular elements surrounding the dead oocyte to swell and migrate into the fibrin, forming either immediately an epithelium-like membrane or powerful cellular streams. The latter may later become distributed in a layer and transform into membranes. At the early stages of cultivation the growth may proceed both in tracts and in membranes. Later on it proceeds prevalently on a plane. The growth in membranes may proceed simultaneously in several planes. In this case the character of the membrane changes according to the conditions of its existence. If the membrane grows in the thick of the fibrin, the cells composing it lie closer to one another, the borders of the membrane break off rather abruptly. They may produce compact outgrowths penetrating into the fibrin.

Quite another character is shown by the more frequently encountered membranes growing on the surface of the fibrin clot. Their cells are greatly flattened out. The borders of the membrane gradually lose their compactness and form outgrowths at its periphery. These outgrowths in some places end by reticular structures and isolated cells of various shapes. The liquefaction gives rise to the formation of «windows», so typical of the mesothelium. The structure of the membranes and the distribution of the elements forming them show a great variability. Frequently one encounters the characteristic arrangement of alternating areas of polygonal cells and areas composed by elongated cells, which has been described for the mesothelium by Khlopin and Mikhailov.

Alongside of membranes with this typical cell distribution, one encounters membranes composed by elongated elements. In the cells it is possible to distinguish an endoplasmatic and an ectoplasmatic part. An intense mitotic multiplication is observed in the membrane cellular elements. The cells in some cases adhere to one another, in others are separated by intercellular spaces, resembling the stomata and stigmata of mesothelial membranes (in preparations treated with silver these spaces stain black). In these cases the cells are united by means of fine protoplasmatic bridges. With the increase of the number and size of these intercellular spaces, the structure of the membranes becomes more and more loose. Finally, thay may turn into flat elongated structures, consisting of cells of an irregular shape united by isolated protoplasmatic filaments of varying length. In the latter case elements of the membrane acquire a likeness to connective tissue cells of the fibroblast type. If the mellowing of the membrane has not proceeded too far, then after the usual addition of fresh medium the cells flatten out again and adhere to one another by their ectoplasmatic margins; in the opposite case the membrane perishes. The trend towards mellowing and towards the formation of reticular structures, the lowered ability to produce closely adhering cellular complexes, is a characteristic peculiarity of the cultures considered in this report.

Besides the reticular mellowing of the membranes, there is observed a complete isolation and separation of their elements. Isolation and separation of individual cells are characteristic of tissues of this type. In the course of this isolation the initially flattened out elements become gradually round and begin to stain more intensely. Multinuclear elements are also encountered both in the membranes and in the cavities of liquefaction. They are apparently formed by means of the fragmentation and mitotic division of the nuclei and of the increase of the protoplasmatic mass.

In the totality of their transformation under the conditions of explantation the follicular and epithelial cells of the ovary of mammalia differ both from the elements of connective tissue and from those of epithelium of the dermal, enteric and nephrous types. They grow outside of the organism according to a peculiar mesothelial or coelom-dermal type, typical of the tegumental cells of the serous envelopes and of the Sertoli elements of the testis, as well as of the myocardium ([7],[8],[5],[10]).

All these tissues are also related to one another by their origin, since they all develop from the splanchnotomes or the coelomic epithelium.

As to the plate of the sexual cells, in these cultures they live only for a short time, without multiplying, and then degenerate. At the early stages of cultivation they may be sometimes passively carried into the zone of growth by the multiplying follicular cells a small distance from the centre of the tissue bit and there degenerate. Their active migration to the zone of growth does not occur.

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HISTOLOGY

GROWTH AND TRANSFORMATIONS OF GASTRIC EPITHELIUM OUTSIDE THE ORGANISM

By A. V. ZUBAREVA

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The purpose of the present work was to study gastric epithelium, as a representative of the enterodermal group of tissues. Simultaneously we intended to establish certain characteristics peculiar to it and distinguishing it from the other varieties of the tissue type mentioned. The literature on the question we are interested in numbers as yet but few studies, made by the method of explantation ([1–7]).

Stomachs of new-born rabbits were used for this study. A bit of the fundus was cut out and rinsed in a jet of sterile physiological saline solution to the amount of 1.5 litre. The cultivation was achieved according to the hanging drop method in a mixture of chicken embryonal extract and heparinized rabbit plasm. The explants were kept under observation for 30 days. A total of 17 series of experiments were arranged including about 300 individual cultures. The examination of control material showed that the thickness of the fundal mucous membrane of a new-born rabbit is one third of that of an adult rabbit.

The mucous membrane shows deep holes and immature glands. In the glands there are two not distinctly differentiated sections. The section opening into the holes corresponds to their future neck and to their supplementary section and consists of elements similar to the epithelial elements of the holes. The deeper section of the glands consists of uniform indifferent elements possessing the shape of a truncated pyramid. The epithelium paving the wall of the stomach is high prismatic, while in the holes it becomes isoprismatic. At the apical end of every cell of the surface epithelium lie mucoid inclusions. On the free apical surface of the stomach the mucoid is distributed in a thin layer covering the entire interior surface of the stomach. Mucoid inclusions are present also in the upper section of the glands, though in much smaller number, while towards the blind end they disappear completely. On fixed preparations mucoid in the lumen of this section appears in the shape of filaments. Both the mucoid inclusions within the cells and the excreted mucoid on the surface and in the lumina of the glands are stained a bright red by Bast's carmine. It is not possible to distinguish the main cells from the reveting ones in the new-born animals by any of the methods employed.

The explant shows two forms of growth: epithelization and plane growth in membranes. After 24 hours there begins a gradual alteration of the explant's relief. Protuberances appear at the surface as the result of an oedema

of the tunica propria. These protuberances appear as blown out spaces between the holes. The blisters grow in size and may merge together. Simultaneously changes take place in the glands as well: their lumina distend greatly and the cells which pave the lumen flatten out. The cells of the terminal sections of the glands, which under usual conditions do not produce mucoid, begin to secrete it. However, the degeneration which begins usually on the 5—6 day leads to the destruction of the gland. The muscular and the reticular tissues also degenerate, and the entire interior part of the explant after prolonged cultivation transforms into detritus, in which one can see isolated elements outwardly resembling macrophages. Epithelization begins early. On the third day already the explant is covered on all sides by newly formed epithelium, mitoses being always visible in the epithelial layer. The new formed epithelial elements contain mucoid inclusions, though in smaller number than in the controls. Formation of mucoid could be observed until the 11th day after explantation. Membranes are formed early on the 2nd day, and it is necessary to note their extremely speedy growth and destruction. They are composed by high polygonal cells with round or oval nuclei. The protoplasm of the cells on fixed preparations is structural around the nuclei and homogeneous at the periphery. In the membranes themselves no mitoses are observed, but they are always visible in the holes. The size of the cells and their nuclei in the membranes increases greatly in the distal direction in respect to the explant. The author was unable to detect clean-cut mucoid inclusions in the elements of the membrane. The phenomena of extensive growth, the epithelization of the surface of the explant, the degeneration of the glands and the oedema of the culture compose one common process, which finally leads to the levelling out of the surface of the explant.

The facts described above allow to draw certain conclusions regarding the place of the epithelium of the stomach in the system of epithelial structures. Gastric epithelium possesses all the distinctive features of the enterodermal group of tissues, i. e.: horizontal anisomorphy, distribution in one layer and in one range, presence of prismatic elements, intercellular bridges (appearing under in vitro conditions) and closing bands. Yet, it shows a number of peculiar characteristics distinguishing it from the enteric epithelium proper. It is known from the works of Khlopin, Fischer, Törö, Khlopkov and Kapel that explants covered entirely with enteric epithelium preserve their viability for a long time and in swelling increase greatly in size, this finding its explanation in the marked absorptive function of enteric epithelium.

Contrariwise, epithelized explants of the stomach wall after a short-lived swelling of isolated areas gradually shrink, strongly decreasing in size, while their contents (muscle, connective tissue and glands) degenerate. This should be accounted for by the fact that under normal conditions there is practically no absorption in the stomach and that the function of the epithelium paving the stomach wall is not absorptive, but protective in the first place. A like function belongs to the mucoid produced by the cells of gastric epithelium.

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MICROBIOLOGY

BIOLOGICAL METHOD OF DISCRIMINATING MICROBACTERIA

By A. I. KORENYAKO

(Communicated by W. L. Komarov, Member of the Academy, 3. III. 1939)

When quantitative and qualitative tests for the presence of bacteria in natural substrata are to be made in great numbers, there is a need of a rapid and easy method for determining bacteria. Unfortunately, however, no such methods are available which would permit us within a short time and with certain ease to discriminate not only the various species, but even many of the genera. Laborious and time consuming is, for instance, the determination of nodule bacteria. And that of microbacteria, which are known (1) to be widely spread in nature, especially in soil, is a matter of still greater difficulty. With regard to a great many species of microbacteria, it is very difficult and sometimes well-nigh impossible to distinguish them from other non-sporiferous bacteria on the basis of morphological and cultural characters. Only when his decisions are based on a quite familiar knowledge and assiduous study of the representatives of that group, can the analyst succeed in this task. As it is not the case with most of the workers, the microbacteria are often confused with other bacteria, and this seems to be a factor in the scarcity of our knowledge about their proper rôle in natural processes.

The every day practice of microbiologists urgently requires such methods to be devised, by whose means a rapid diagnosis of the various groups of bacteria would become possible.

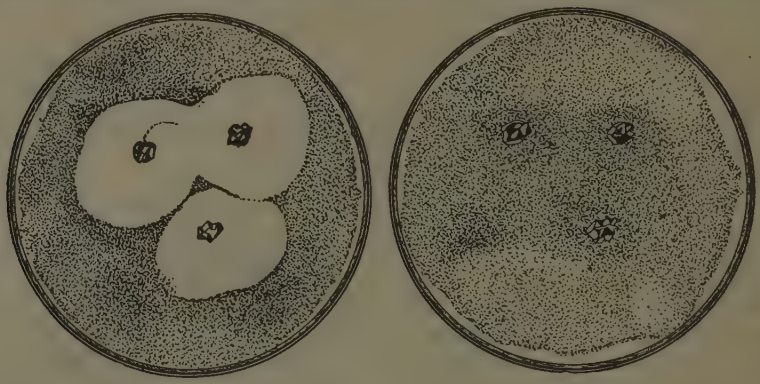
In the present work I am suggesting a method which will afford the possibility of readily discerning the microbacteria from other non-sporiferous bacteria, in general, and from nodule bacteria, in particular, since this latter type is often encountered in practice and most difficult to distinguish from microbacteria. The method is based on the properties of some actinomycetes, so-called antagonists, which check the development of certain species of bacteria (sporiferous and oth.), among them number of microbacteria, without suppressing in any way the other, non-sporiferous species, including nodule bacteria. At our Institute we have studied about 500 strains of microbacteria from different soils of this country.

By acting with cultures or filtrates of actinomycetes upon microbacteria, we have been successful in establishing that the latter are always depressed by actinomycetes-antagonists, while all the other non-sporiferous bacteria, not excepting the nodule bacteria, live in harmony with them. This circumstance made it possible to recur to a simple and easy method for distinguishing microbacteria from other non-sporiferous bacteria.

The method consists in the following: the culture of non-sporogenous bacilla under test is transferred to the respective nutrient medium (MPA, bean-agar, čapek-agar etc.) in Petri dishes and spread with a spatula carefully over the whole of the agar surface. Pieces of a grown actinomycetis-

antagonist are then placed upon, and the dishes transferred to the thermostat. The result is obtained after one or two days, according to the rate of growth of the organism tested.

In the case of microbacteria a sterile zone, 10—30 mm and more, is formed around the actinomycetis, while no zone whatever is observed in the case of other non-sporiferous bacteria, as for instance, the nodule bacterium or one of the group *Pseudomonas*. The action of actinomycetis on micro-



bacteria and its action on nodule bacteria is shown in the figure. In our experiments we used the strain *Act. violaceus* 41 as antagonist, but many other actinomycetes-antagonists are also fit for that purpose. It is a simple matter to establish whether the strain concerned is an antagonist, and hence may be used for the determination of bacteria, or otherwise. One need only to inoculate it simultaneously with, say, *B. mycoides*, and if antagonist it will call forth the formation of a sterile zone within the continuous growth of *B. mycoides*. A more pronounced antagonistic effect of actinomycetes is obtained by growing them in synthetic medium of the following composition: distilled water, 1000 cm^3; KCl—0.5 g; $MgSO_4$—0.5 g; K_2HPO_4—1.0 g; $NaNO_3$—2.0 g; $FeSO_4$—0.01 g; glucose—30.0 g; agar-agar—20 g; pH—7.2. Actinomycetes-antagonists grown in protein medium (MPA, MPG etc.) are but slightly active if at all.

The method suggested has been verified on various material within a period of 2 years. More than 1000 strains of non-sporogenous organisms from different soils of this country and 25 strains isolated from wheat grain have been investigated. The study involved moreover about 100 strains of nodule bacteria from peas, kidney-beans, lucerne, soya, vetchling, serradella, horse-beans, spring vetchling, melilot. The species *Rh. trifolii* alone was represented by 50 strains from various soils of the Soviet Union. And in all cases without exception have we obtained the same result. All microorganisms belonging to microbacteria invariably formed a zone around the actinomycetes, while not a single case was observed in which non-sporogenous or nodule bacteria had formed such a zone; on the contrary, they developed in harmony with the actinomycetes.

We are therefore entitled to say that the method described may safely be used for the purpose of rapidly discriminating microbacteria from non-sporiferous, especially nodule bacteria.

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GENETICS

NEW DATA ON SPONTANEOUS MUTATIONS

By J. M. OLENOV

(Communicated by N. I. Vavilov, Member of the Academy, 25. II. 1939)

It is generally accepted among geneticists that spontaneous mutations originate autonomously, independently of the development of the individuum, and that it is possible to speak of periods that separate one mutation of the gene from another without taking into consideration the separate individual cycles. Such is the point of view advanced in most of the papers in which the question is treated under its general aspect. However, our opinion is that although the experimental study of the problem is as yet (leaving aside the Lamarkist experiments) at its initial stages*, even on the ground of a priori considerations it is impossible not to question the correctness of the generally accepted solution.

Among the conditions that are of decisive importance in the development of the individual, without directly affecting its germ-cells, the nutritive regime must be mentioned in the first place. The purpose we had in view was to study the rate of the mutation process in *Drosophila melanogaster* under conditions of an avitaminous B_2 regime.

For our investigation we made use of the line *white*. By way of manifold passage of the flies (Guyénot method, 1911), a complete elimination of the microflora was achieved and an aseptique line obtained.

It was found as a result of a series of experiments that under conditions of an avitaminous B_2 regime more than 99% of individuals died without reaching the pupal stage. About a third of larvae reached the age from 18 to 22 days (the experiments were carried out at 25°C; in control cultures on a sterile yeast medium the flies began to emerge on the 10th day).

If into a test tube containing 20 day old larvae yeast medium is added, the larvae reach the pupal stage after a period of 5 to 6 days and the flies begin to emerge on the 10th day, that is 30 days after the eggs had been laid. Larvae that have lived be it only 24 (or even 16 to 18) hours, after hatching on a sterile yeast medium, on being subsequently placed in an avitaminous B_2 medium, go over to the pupal stage and the flies emerge (at the same time as the control individuals) on the 10th day, the imago stage being reached in nearly 100% of cases.

Thus, 3 groups of individuals could be employed for genetical purposes: 1) the control individuals which developed on a yeast medium; 2) individuals given yeast medium on the 20th day of their larval life, 3) individuals which had fed on yeast medium for a period at the utmost of 24 hours after

* Schlösser, Döring.

hatching and continued their development on an avitaminous B_2 medium. The individuals of the 2nd group, which received the vitamine B_2 although at a very late period but in sufficient amounts, did not differ from the controls in regard of the duration of life and fertility (males only were studied). It stands to reason that these specimens at the period preceding the addition of yeast medium have been strongly affected by the avitaminous B_2 regime. In individuals of the 3rd group the duration of life is greatly diminished; about half of the males of this group are infertile in the first four days following their emergence. It is obvious that the vitamine B_2 does not only influence metamorphosis (minimum amounts suffice for securing the setting in of the pupal stage and the emerging of the fly), the organism as a whole being likewise strongly influenced.

The spontaneous mutational process in the *X*-chromosome of the control ♂♂ and the ♂♂ of the second group was analyzed by the *ClB* method. Experimental and control ♂♂ were crossed with ♀♀ $\frac{Clb}{sc\ v\ f\ car}$ on the day of the flies emerging from the pupae. The parents were taken from the test tubes two days later. Such mutations as caused a 100% mortality in F_2 were alone taken into consideration. All the mutations were verified in F_3.

The rate of mutations found in the experiment and in the control (Table 1) is about the same in both cases; it is somewhat lower in the experiment, but the difference is quite immaterial. Diff.$=0.06\pm0.11$. At the same time if the number of originated mutations corresponded to the time elapsed, the percentage of mutations brought into evidence in the experimental material (duration of development 30 days) should have been three times as large as in the control (10 day duration of development).

Table 1

Lethal Mutations in the *X*-chromosomes of Control ♂♂ and Experimental ♂♂ of the 2nd Group

Control		Experiment	
Number of chromosomes studied	Number of lethal mutations	Number of chromosomes studied	Number of lethal mutations
352	1	730	1
676	1	741	1
489	1	944	2
880	2	817	1
2397	5 %$=0.21\pm0.08$	3232	5 %$=0.15\pm0.07$

Diff. $= 0.06 \pm 0.11$

One may, however, assume that the number of originated mutations for unit time is actually the same in the experiment as in the control, and that the result obtained is due to germinal selection more intensive in individuals of the 2nd group than in the control ♂♂. Therefore, in order to verify the validity in this case of the germinal selection hypothesis, and at the same time with the view of completing the data already obtained, the spontaneous mutation process was studied in the II chromosome. We again employed for our analysis control ♂♂ and ♂♂ of the 2nd group. The analysis was carried out according to the $\frac{Cy\ sp}{L^2}$ method and such mutations only were taken into

account, which in F_3 produced death of all individuals homozygous for the corresponding mutation. All mutations were verified in F_4*.

The percentage of lethal mutations found in the II chromosomes in experimental ♂♂ of the 2nd group is about on a par with the rate of mutations in the II chromosomes in the control or even somewhat lower than the latter (Table 2). From the statistical point of view the difference is not quite real. Diff. $= 0.41 \pm 0.21$. By carrying out the necessary computations, proof may be obtained that the number of mutations originated in the experiments per unit time differs from what we found in the control. The number of mutations found in the control (M^c) $= 42 \pm 6.45$. Hence, the minimum number of mutations possible in the control ($M^{c\ min}$) $= 42 - 6.45 \times 3 = 22.65$. The next question is: which would be the lowest of all possible numbers in the experiment ($M^{e\ probable\ min}$), if in experimental ♂♂ the rapidity of the mutation process was the same as in the controls? Obviously, it amounts to the minimum number of mutations in the control ($M^{c\ min}$) multiplied by 3 (the development of the flies in the 2nd group is three times as long as that of the control flies) and further multiplied by 1.04 (in the experiment the number of chromosomes investigated was somewhat higher than in the control). Thus, $M^{e\ probable\ min} = 22.65 \times 3 \times 1.04 = 70.07 \pm 8.30$. The number of mutations actually found in the II chromosomes in experimental ♂♂ (M^e) $= 27$. It follows that statistically the difference between the number of mutations found and the number of mutations to be expected in the case if the rate of the mutational process did not depend on the condition of the individuals development, is quite real. Diff. $= 43.07 \pm 8.30$.

Table 2

Lethal Mutations in the II Chromosomes of Control ♂♂ and Experimental ♂♂ of the 2nd Group

Control		Experiment	
Number of chromosomes studied	Number of lethal mutations	Number of chromosomes studied	Number of lethal mutations
832	9	648	4
695	7	837	5
938	10	869	6
720	6	903	6
239	4	338	3
511	6	494	3
3935	42 %=1.07±0.16	4089	27 %=0.66±0.13

Diff. $= 0.41 \pm 0.21$

Note: 1) $M^c = 42 \pm 6.45$. 2) $M^{c\ min} = 42 - 6.45 \times 3 = 22.65$.
3) $M^{e\ probable\ min} = 22.65 \times 3 \times 1.04 = 70.07 \pm 8.30$. 4) $M^e = 27$
Diff. $_{3-4} = 43.07 \pm 8.30$.

* The gene *sp* permitting of a verification of the cultures in which single non-*Cy* flies were found, evidence was obtained in this manner that the appearance of such flies may be due to a crossing over between the *Cy sp* chromosome and the chromosome bearing the newly originated lethal. The percentage of crossovers is always below unity. Such lines were, of course, taken into consideration as possessing a lethal mutation.

The experiment just described contains material from which it may be seen that the assumption of a germinal selection acting in our case has no validity. The lethal mutations originated in the II chromosome, different from the sex-linked mutations, remain heterozygous up to the reduction division, so that no distinct decrease in the numbers of mutations found can be caused by germinal selection. At the same time the relation of the rate of mutations found in the experiment to that in the control is about the same for the *X*- as for the II chromosome. A decrease in the number of mutations found in the experiment as compared with the control (in so far it is possible to speak of such a decrease) is even more strongly expressed in the II chromosome (compare Tables 1 and 2).

In the data given above yet another fact may be noted which seems to be of great importance, viz. whether it is a mere coincidence that the percentage of mutations relative to the number of chromosomes studied proves to be about equal in ♂♂ of the 2nd group and in control ♂♂? Does not this fact indicate that in ontogenesis causes are operative that produce accumulation of a certain number of mutations in the individual's germ-cells at the time of emergence of the flies? To solve this question individuals of the 3rd group were employed. Their development lasts 10 days, that is, it does not differ from the duration of the development of the control group. Therefore from the study of the mutation process in ♂♂ of the 3rd group, it is possible to decide whether the decrease in the rapidity of the mutation process found in the previous experiments is due to a more prolonged development of flies of the 2nd group, or, else, it is due to some other causes (say, to the direct effect of a decrease in the intensity of metabolism).

For the analysis only such males belonging to the 3rd group were employed as were infertile in the first 4 days after emergence of the flies, i. e. such specimens which had most suffered from the infavourable effect of B_2 avitaminous regime. Control males were likewise employed on the 5th day after emergence. As in the previous series of experiments, analysis of the mutation process in the II chromosome was carried out according to the $\frac{Cy\ sp}{L^2}$ method. The method of evaluating the mutation originated was unchanged.

The percentage of lethal mutations found in the II chromosome in the 3rd group proved to equal the per cent of mutations in the control (Table 3). Thus, we may consider as proved that the decrease in rapidity in the mutational process, brought into evidence in the preceding experiments, in which ♂♂ of 2nd group were studied, is due to the fact that in the group under discussion the development is of longer duration. The decisive rôle may quite possibly belong here to a slowing down of the process of spermatogenesis, or, more exactly, in connexion with the latter, to a slowing down of the rate of reproduction («multiplication») of the genes.

Table 3

Lethal Mutations in the II Chromosomes of Control ♂♂ and Experimental ♂♂ of the 3rd Group

Control			Experiment		
Number of chromosomes studied	Number of lethal mutations	% of lethal mutations	Number of chromosomes studied	Number of lethal mutations	% of lethal mutations
941	7	0.74	688	5	0.73

Diff. = 0.01 + 0.43

The results obtained in the present paper enable us to conclude that the number of mutations formed in an individual's germ-cells up to the close of metemorphosis is characteristic of the corresponding period of life of the individual. In our opinion this fact proves the erroneousness of the current view that the spontaneous mutation process is autonomous, i. e. independent of the development of the individual.

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PLANT PHYSIOLOGY

WAVE LENGTH OF LIGHT APPLIED DURING FIRST STAGE OF DEVELOPMENT AS A FACTOR IN FURTHER GROWTH AND DEVELOPMENT OF PLANTS

By A. A. KUZMENKO

(Communicated by A. N. Bach, Member of the Academy, 25. II. 1939)

Our experiments on irradiation of germinating seed with a light of varying wave length have shown this treatment to produce a strong influence upon the further growth and development of plants, proceeding under natural conditions of illumination (1). In the experiments the growth of the germinating seed under irradiation was checked by limiting the water supply. The results obtained from a 10—14-hour irradiation varied according to the plant tested. In general, most of the plants under experiment accelerated their growth and development considerably in response to the irradiation of their seed with the long-wave part of the spectrum (red-yellow light). This effect was most conspicuous in the case of *Perilla ocymoides* and *Tr. vulgare*(2). And no less strongly was it expressed in some varieties of tobacco and makhorka (*Nicotiana rustica*) subjected to 10-hour irradiation (3).

These results prompted us to proceed in our investigation with the aim to throw light on some questions arisen in connexion with the first experiments.

Among the objects we had before us was to establish how various plants respond to the factor of wave length when they are passing the light stage. That it is a very important point to clear up has already been emphasized by us in discussing the results of our experiments of 1936 (3). In 1938 we found an opportunity to start with some experiments in that direction and it is the results obtained which we are going to set forth here.

The experiments involved the following plants: tobacco makhorka, tomatoes (three varieties), cabbage, strawberry, camphor-basil, arachis, etc. The seedlings were grown in usual hotbeds being given light of varying composition. The experiments were run in the new botanical gardens of the Academy of Sciences of the Ukrainian SSR.

Acetyl-cellulose film was made use of for filtering sunlight, since in many important respects it is more suitable for the purpose than glass(4). The filters were made upon our behalf at the Physico-Agronomical Institute in Leningrad, their manufacturing being guided by D. V. Fedorov. They were fastened in position on hotbed frames which were specially adjusted to carry them. We had at our disposal four kinds of coloured filters (orange, blue, red and green) and a transparent colourless film. Most of the experiments were run with the first two colours and the transparent film, which served

as control. On fastening them in position, we strictly followed all the items of a special instruction.

The filters transmitted the following portions of the visible spectrum.

1. Transparent—all the visible rays of the spectrum.
2. Orange—rays ranging in wave length between 750 and 575 mμ.
3. Blue—rays with λ equal about 500—400 mμ and a narrow red band from the extreme part of the spectrum.

The experimental plants experienced the action of coloured light during 13—20—26 and more days after seedlings had appeared above ground. Then they were pricked off into vegetation vessels or in a plot and grew under natural daylight until complete ripening. All plants were cared for according to their respective requirements, the conditions being the same in each of the variants within the crop concerned. Exact observations were made on their development and dynamics of growth. Some of them were subjected to chemical analysis for the content of carbohydrates, organic acids, essential

Fig. 1. *Ipomaea purpurea*, 13 days after seedlings appeared above ground: *А*—under transparent film, *Д*—filter II, *К*—filter III.

oils and alkaloids. Chlorophyll was estimated in the leaves of tobacco plants towards the end of the period under filters.

Ipomaea purpurea. This decorative plant was kept under filters for a short time—during 13 days after the seedlings pushed out of the ground. Fig. 1 gives an idea of the growth of this plant in response to light of various colours. The differences the plants showed under filters were preserved, when they were growing under natural daylight after being transplanted in a garden bed. Besides their lagging behind in growth, the plants under filter III delayed flowering for 4 days compared to plants under filter II. In other words, the elimination of long-wave part of the spectrum for 13 days following the appearance of seedlings slowed down the subsequent development of plants.

T o m a t o e s. Three varieties of tomato have been tested: Humbert, Erliana № 12 and John-Bair. The young seedlings were kept under filters for 26 days and then pricked off into vegetation vessels or in field ground, where they were growing under natural daylight. For all of the varieties tested a great difference was observed between the plants from under filters II and III. The latter lagged behind in growth and their leaves were not so intensely coloured as those of the former (the concentration of chlorophyll in the plastids was lower). The difference would not disappear until they had finished ripening. Fig. 2 is a photograph taken from experimental plants of one of the varieties tested 20 days after they had been pricked off into vessels. The retardation of growth and development of vegetative organs did not fail to affect both flowering and fruiting. The plants from under filter III markedly delayed the outset of flowering and fruiting, and this of course remained not without influence on the quality and amount of yield. The chemical analysis of the latter revealed notable differences between the

varieties, as regards their response to the treatment of seedlings with coloured light.

Carrot. The maximum length of the period these plants remained under filters was 44 days. The effect of the treatment was to induce a rather sharp difference, especially in the variants I and II, on the one side, and in variant III, on the other. The difference continued throughout the subsequent development of plants and it told upon the yield of root and above ground organ. This is shown in a representative manner in Fig. 3. Microscopical investigation of

Fig. 2. Tomatoes, variety John-Bair, treated for 26 days following the appearance of seedlings: *1*—under transparent film, *2*—filter II, *3*—filter III.

Fig. 3. Carrot, 44 day treatment: *1*—under transparent film, *2*—filter II, *3*—filter III.

plastids in harvested root of carrot has revealed an increased concentration of carotin in the cortical part of the root of plants from under filter II.

Strawberry hybrid № 568-36, variety «Umanskaya». It is an interspecific hybrid produced by crossing garden and wild strawberry

Fig. 4. Strawberry hybrid № 568-36: *1*—under transparent film, *2*—filter II, *3*—filter III.

(*Fragaria grandiflora* × *Fragaria Elatior*) by S. Kh. Duka, in 1936. Vegetatively propagated material for the experiment was supplied early in spring by its breeder from Uman. Two sets of tests have been carried on with this plant. In the first set the treatment with coloured light lasted from April 19 till May 10; in the second set, from May 11 till June 2. The rest of the time the plants were growing under natural daylight.

The outset of flowering, formation and ripening of fruit were rather considerably accelerated in the plants from under filter II, especially in the first set of tests. The yield of fruit increased, too. This plant proved to be very plastic in response to the treatment in question and the experiments with it are being continued.

To a certain extent the data obtained are in agreement with what has been stated by V. P. Malchevsky ([5]) and N.N. Gortikova([6]). The method of growing seedlings under coloured light can, after being tested and developed in more detail, be applied to practical socialist plant breeding work. The influence of light composition upon various physiological processes in plants is worth close systematical study.

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EXPERIMENTAL ZOOLOGY

EXPERIMENTS ON TOTAL EXTIRPATION OF THE WHOLE ENTODERM IN TRITON EMBRYOS

By B. I. BALINSKY

(Communicated by I. I. Schmalhausen, Member of the Academy, 6. III. 1939)

The significance of a certain embryonic part for the development of other parts of an embryo may be investigated by extirpating the part in question at a sufficiently early stage of development. This method of investigation was applied, for instance, when the rôle of the auditory labyrinth in the formation of the cartilaginous capsule was shown by extirpating the ear vesicle (experiments by Filatow, Sternberg, Luther). It is much more difficult to extirpate greater parts of the embryo, especially to totally extirpate a certain germinal layer, as for instance the whole entoderm. Such an experiment would be of great interest, as it would be possible in this way to explore the rôle of the entoderm in the development of the whole embryo. As a matter of fact, the extirpation of the whole entoderm proved to be practicable. Mangold(1) was the first to succeed in this experiment. It is a matter for regret that Mangold published only a short preliminary note on his experiments; the entodermless embryos, as may be concluded from that note, were not investigated in sections.

I obtained several entodermless embryos during my experiments in spring 1938, the primary aim of which was to investigate the determination of entodermal organs in amphibian embryos. I extirpated the whole entoderm of *Triton taeniatus* embryos at the stage of neurulation. In the youngest embryos operated the medullary plate had just begun to be visible, in the oldest—the medullary folds had already united (Harrison's stages 14—19). Having made a long incision on the ventral surface of the embryo, I began to separate the ecto- and mesoderm from the entoderm on both sides and eventually extracted the entoderm as a whole mass. The separation of the entoderm from ecto- and mesoderm is easily attained in these stages, only in the prechordal and blastoporic regions the entoderm has to be cut off from the remaining parts. The ecto- and mesodermal part of the embryo are originally flattened out, but later on they roll together, the margines of the cut ectoderm unite, and there are again formed whole embryos that consist now only of ectoderm and mesoderm (including notochord). I have examined in sections 7 such entodermless embryos, that survived from 8 to 14 days after the operation. The following is a short report of this examination (compare also the graphical reconstruction of a typical case, presented in the figure). The examination of sections shows that the entoderm was extirpated very accurately, in several embryos there is not a single frag-

ment of entodermal tissue. The notochord is normally differentiated, as could be expected, but its shape is distorted, the notochord has not elongated normally, probably, as a result of purely mechanical disturbances. The dorsal muscles are more or less normally differentiated. The pronephroi are normally differentiated, but violently displaced from their normal position. In embryos that lived long enough the fore-limb rudiments are present. Apparently, the pronephros and fore-limb, though normally adjoining the entoderm, may develop independently of the latter. The coeloms in all entodermless embryos are very much bloated, thus giving the entodermless embryos a very characteristic appearance: a narrow axial part of the body is wedged between two thin-walled vesicles (the right and left coelom, as a rule, remain separated from one another). The bloating of the coeloms occurs in the neck region of the embryo. The rudiments of pronephroi and limbs are thus pushed far back, and lie on the posterodorsal surface of the vesicular coeloms. The nervous system of entodermless embryos is comparatively normal, only in the anterior brain region there are synophtalmic or even cyclopic defects. This is a result of the extirpation of the utmost anterior part of the archenteric roof, the presence of which, as was shown by Adelmann and others, is indispensable for the normal development of the fore-brain and eyes([2],[1],[3]). The absence of entoderm produces the greatest effect in the development of the branchial region, as was also observed by Mangold([1]). The entodermless embryos possess no external gills, no cartilaginous branchial skeleton, no visceral arches at all. The absence of external gills is in correspondence with the data by Harrisson([4]), Severinghaus([5]) and Ichikawa([6]), who showed that these organs develop in result of an induction on the part of the branchial entoderm. In some of the entodermless embryos I have found in sections an ectodermal stomodaeal invagination, that developed in spite of the absence of mouth entoderm. I will not discuss this fact here, as it will be the subject of another publication. When the stomodaeal invagination is present, there are formed in its proximity small cartilages, that represent a very rudimentary skeleton of the anterior part of the head. In other entodermless embryos there are no skeletal parts whatever. Finally a characteristic feature of the entodermless embryos is a total absence of the heart in all such embryos. The absence of the heart was first noted in living embryos, where even a small fragment of cardial tissue would have been made conspicuous by its pulsation (having very little yolk, the entodermless embryos are very transparent). The absence of the heart was later confirmed during the microscopic examination of sectioned entodermless embryos. The blood rudiment in these embryos is also absent. The absence of the heart in entodermless embryos is noteworthy, since in very many investigations it is assumed that the heart develops from the lateral mesoderm, and consequently the rudiment of this organ might not have been extirpated in my expe-

Graphical reconstruction of entodermless embryo (ventral view): *M*—central nervous system, *Oc*—eyes, *Olf*—olfactory rudiment, *Ch*—notochord, *Mc*—muscles, *Pn*—pronephros, *E*—fore-limb rudiment, *St*—stomodaeal invagination, *C*—cartilages.

riments. It is true, that a ventral incision could have considerably disturbed the heart rudiment, and yet it is known that the heart rudiment possesses a high regulative ability, and in Ekman's experiments(7) the hearts developed normally after the extirpation of a broad median strip of the heart rudiment. It should be assumed that in the neurula stage the entoderm possesses something that is of importance for heart development. Apparently the problem of heart development in amphibians should be reinvestigated.

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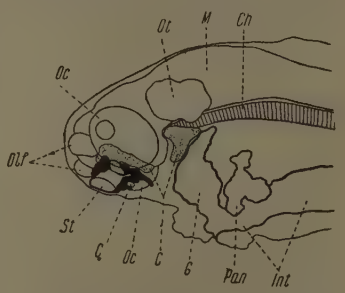
EXPERIMENTAL ZOOLOGY

ON THE DEVELOPMENT OF THE ECTODERMAL STOMODAEAL INVAGINATION IN AMPHIBIANS

By B. I. BALINSKY

(Communicated by I. I. Schmalhausen, Member of the Academy, 6. III. 1939)

The development of the stomodaeal invagination was recently a subject of experimental investigation by several authors, who worked with embryos of different amphibians ([1–7]). Unfortunately, there exist very considerable discrepancies between the data of the above-mentioned authors, whose opinions differ in particular as to the time and origin of determination of the stomodaeal ectoderm. Miss Adams ([1,2]) extirpated and transplanted the mouth ectoderm in *Amblystoma punctatum* embryos and also extirpated the entoderm of the mouth region in the same species. This author found that ectodermal invagination is induced by the oral entoderm in the young tail bud stage, and that from this stage the stomodaeal entoderm acquires the ability for independent formation of rudimentary ectodermal invaginations, without a cavity but developing teeth. Marcus ([3–6]) carried out on embryos of anuran amphibians experiments similar to those of Miss Adams; he found, however, that in anurans the ectodermal stomodaeal invagination is determined as early as in the neurula stage and may develop more or less normally in the absence of connexion with the oral entoderm. According to Marcus, in the neurula stage the oral entoderm is separated from the mouth ectoderm by a layer of mesoderm; consequently, it is impossible to suggest the existence of an induction of the ectodermal stomodaeal invagination by the oral entoderm. Finally, Ströer concludes from his experiments, on the details of which we need not dwell here, that the ectodermal mouth invagination is induced by the anterior wall of the fore-gut between the middle gastrula and the neurula stages; yet, according to Ströer, the ectodermal mouth invagination even after this stage may differentiate only if there is a constant connexion with the mouth entoderm.



Graphical reconstruction of head in case Uen-64 (lateral view): *M*—central nervous system, *Ot*—ear vesicle, *Oc*—eye, *Olf*—olfactory rudiment, *Ch*—notochord, C—cartilages, *St*—stomodaeal invagination, *G*—stomach, *Pan*—pancreas, *Int*—intestine.

Some experiments which I made during the spring seasons of 1937 and 1938 may serve to clear up the problem in question. My experiments consisted in the extirpation of more or less considerable parts of the entoderm in embryos of different stages, beginning with the blastula stage and ending with the optic vesicle stage (Harrison's stages 9—21). The entoderm was extirpated either totally ([8]) or only the anterior entoderm was removed, i. e. oral and pharyngeal entoderm. Eventually in a number of embryos there was no contact between the entodermal tissue and that part of the dermal ectoderm, where the stomodaeal invagination is formed. In several further cases there was no contact between the ectoderm and entoderm of the mouth region in result of a disturbance of gastrulation. Already in living embryos it was seen that some of them possessed ectodermal stomodaeal invaginations. An examination of the operated embryos in sections confirmed this fact (figure). At the same time it was found that the stage, at which the operation was made, was of great importance for the result of the experiment.

Development of Ectodermal Stomodaeal Invagination in Absence of Contact with Oral Entoderm

Stage of operation	Stomodaeal invagination absent	Stomodaeal invagination present		
		no external opening	narrow cavity and small opening	large cavity and broad opening (mouth comparatively normal)
9	Uen-35, Gen-109, Gen-214, Uen-46, Uen-47, Uen-50	—	—	—
10	Uen-4, Uen-7, Uen-13, Uen-25, Uen-30, Uen-43, Uen-53	Uen-9	—	—
11	Uen-21	—	—	—
12	—	—	—	—
13	—	—	—	—
14	Ent-376, Uen-66, Uen-70, Uen-72, Uen-74	Uen-69	Ent-468	Uen-64
15	Ent-377, Ent-441, Uen-71	—	Uen-55, Uen-60, Uen-62	—
16	Uen-58, Uen-75	Ent-457	—	—
17	Uen-57, Uen-76	—	—	—
18	Uen-77	—	Uen-67	Uen-65
19	—	Uen-59	—	Uen-54
20	—	—	—	—
21	—	—	—	Uen-61, Uen-68

The table shows the distribution of the operated embryos according to stage of operation, and results that were obtained in each case. We may distinguish three phases in the development of the stomodaeal invagination, that differ in the degree of dependence on the entoderm. If the entoderm was removed before gastrulation, or during the beginning of gastrulation (stages 9—10), that is before a contact was established between the entoderm and ectoderm of the mouth region, then the stomodaeal invagination was not formed. Only in one case (Uen-9) a small stomodaeal invagination developed after an operation in stage 10. But in the record of this case there is a note, that the archenteron was considerably developed and was already expanded at its anterior end. It is consequently probable that the entoderm in this case had already been in contact with the future mouth ectoderm.

at least for a short time; this contact seems to have been sufficient for the production of a small inductive effect. That the presumptive mouth ectoderm may easily react to weak inductive stimulus is proved by cases, in which a contact between the ectoderm and entoderm was established to a very small extent, and in spite of this there developed sometimes a large ectodermal invagination. If the entoderm was removed in the neurula stage (stages 14—17), i. e. after the mouth ectoderm was for some time in contact with the entodermal fore-gut, an autonomous development of the ectodermal stomodaeal invagination occurred in a number of cases, and it may be even of considerable size (comp. case Uen-64, figure). And still the determination of the stomodaeal invagination is far from being complete in these stages: in a part of the embryos the ectodermal mouth failed to appear altogether, and even where an invagination is formed, it is mostly poorly developed. Obviously, the entoderm in these cases continues to produce its influence on the mouth ectoderm and this influence is still important for the development of the latter. In the late neurula stage (closure of medullary folds) and in the optic vesicle stage (stages 18—21) the ectodermal stomodaeal invagination may develop autonomously even without further contact with the oral entoderm, and this development may be rather complete. In a few cases the ectodermal invagination differs but little in dimensions, form and histological differentiation from a normal one. The development of the stomodaeal invagination approaches the type of self-differentiation, at least in respect of the entoderm, as there are established intimate relations between the stomodaeum and other cephalic organs (the olfactory organs, the skeleton). But even in these late stages a strong disturbance may apparently suppress the development of an ectodermal mouth. Among the cases at my disposal such a suppression of the mouth development occurred but once (Uen-77), and even this case was not quite reliable: there is a note in the record that a piece of the ectoderm was unintentedly torn away in this case. This bit might have included oral ectoderm. My results show a way to conciliate and generalize the apparently contradictory data of the investigators mentioned in the beginning of this article. First of all, it must be stated that there is no fundamental difference between the *Anura* and *Urodela* in the development of the ectodermal stomodaeum. Both in the *Urodela* (my data) and in the *Anura* (Marcus' data) the ectoderm possesses an ability to develop a stomodaeal invagination in the absence of entoderm as early as in the neurula stage. Marcus believes that this ability of the ectoderm to develop stomodaeal invaginations originates quite independently of the entoderm. But this opinion is based solely on the assumption that in the neurula the entoderm and ectoderm in the mouth region are separated by a layer of mesoderm, and that their contact is established later on. The assumption made by Marcus is highly improbable and is in direct contradiction to Vogt's data on the gastrulation of anuran amphibians (⁹). It seems obvious that in *Anura* as well as in *Urodela* a contact between the ectoderm and entoderm in the mouth region is established in the late gastrula stage. It is in this stage that the ectodermal stomodaeal invagination is induced by the oral entoderm. From this same stage also the stomodaeal ectoderm acquires some ability for autonomous differentation and this ability is increased in the following stages. The fact that Ströer failed to observe an ability of the mouth ectoderm for autonomous differentiation and that Miss Adams found it only at later stages and in a small degree—should be accounted for by a violent disturbance of the ectodermal material in their experiments. Miss Adams removed the oral entoderm through an incision in the middle of the mouth region; Ströer investigated the capacity of ectoderm for self-differentiation only in transplantation experiments, in which the ectoderm

must have been both disturbed and, perhaps, exposed to inductive influences in its new position. I do not wish to say that there are no differences between the development of the stomodaeal invagination in *Anura* and *Urodela*, but it seems that these differences are not fundamental but of a quantitative nature, and consist mainly in a different rate of the stabilization of the determination of this organ.

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DEVELOPMENT MECHANICS

REGIONAL COLORATION AND ITS DETERMINING FACTORS IN BIRDS

By W. TH. LARIONOV

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A number of questions concerning the mechanics of development is connected with the variety of the coloration characters in birds. These are, first, the factors responsible for the changes of the type of the plumage in time, the problem of age and seasonal dimorphism, secondly, the factors of sexual dimorphism and, thirdly, the cause of the distribution of coloration over the various parts of the bird's body.

At present, there are no doubts that the three mentioned types of dimorphism are hormonally determined. Thus, the sexual differentiations develop under the influence of the ovary hormone, they disappear in females after castration and are restored after reimplantation, while in males they are induced by the female sexual hormone. Both age and seasonal changes in the type of the plumage are also controlled by the changes in the type or intensity of the hormonal functions.

Contrary to that, the conditions determining the difference in the coloration of the plumage on various parts of the body remain unknown up to now. We do not know whether the regional coloration is determined by hormonal factors, or else it is controlled by the factors of the feather anlage itself or, finally, it is due to the interference of the hormonal and local factors.

This problem can be solved experimentally by way of disarranging the regional coloration and inducing generalized colouring instead.

As early as in 1923 Horning and Torrey ([1]) pointed to the fact that in Leghorn cocks subjected to the action of the thyroid hormone the red pigment on the feathers of the back was to some extent changed to black, while the black colour on the lower side of the body remained unchanged. Later on this fact was confirmed by a number of authors ([2—13]).

Basing on the obtained feathers having some common traits with those of a hen, Horning and Torrey, as well as the majority of authors, interpreted their results as an inversion of the sexual characters of the plumage. A few authors rejected this wide spread theory, among them Greenwood and Blyth, who pointed to the incomplete effect of feminization and, in connexion with that, to the possibility to interpret the pigment reaction in another way. On the basis of parallel experiments with the extirpation of the thyroid gland, Greenwood and Blyth came to the conclusion that during hypo- and hyperthyroidism a general change in the coloratoin of the male takes place. From a combined red-black it turns to a certain extent into completely

red (hypothyroidism), or completely black (hyperthyroidism). Such an interpretation being adopted, the results would be referred to the problem of regional coloration instead of sexual dimorphism. However, it did not happen and the problem remained unsolved, the more so, as alongside of those opinions such were advanced, according to which the changes in the coloration during hyperthyroidism reflect either age differences ([14]) or non-specific aberrations of the albinism or melanism type ([13]).

As Greenwood and Blyth disposed of scanty material (a single thyroidectomized specimen), the recent work of Parkes and Seleye([15]) proved of decisive importance. In their investigations, Greenwood's data were confirmed on a great number of specimens. In the experiments of Parkes the black colour of the lower side of the body was changed to red in thyroidectomized Brown Leghorns and to white in Silver Dorkings, the red and the white colour of the upper side of the body remaining respectively unchanged. Still earlier Hutt ([16]) had noticed an opposite effect during hyperthyroidism, namely, the development of black colour instead of the white one (Silver Wyandottes). He interpreted this fact, however, as an index of partial feminization.

In experiments parallel to those performed on hens, Parkes and Seleye observed that the thyroidectomized drake, instead of black colour on its back, developed a rippled pattern which is normally present on the lower side of the body. According to the data by Streich and Swetozarov obtained in our laboratory ([17]), a contrary effect is produced by hyperthyroidism; black coloration on the lower side of the body. All these data were fully confirmed on ducks in new experiments of our laboratory (unpublished).

It seems to us that the material accumulated up to the present time, referring to problem of regional coloration, is sufficient for certain though preliminary generalization.

As far as, on the one hand, under the influence of the hormonal factor, the thyroid gland in our case, the regional coloration is obliterated and its generalization takes place, one should assume that the hormones play a definite rôle in the differentiation of variously coloured parts.

On the other hand, as far as with the decrease of the inductor's activity (hypothyroidism) generalization proceeds at the expense of one of the existing colours (white or red in hens, ripple pattern in ducks), while the increase of the activity proceeds at the expense of another, one should assume that the result of the hormonal action depends on the susceptibility of the various parts. In other words, in that case we must apply the principle of «differential thresholds», previously used by Pézard and M. Zawadowsky in the analysis of sexual characters. Then, as far as the character of the reaction to the change in the activity of the inductor varies in form with different genotypes, though they have some common features (the increase and the decrease of pigmentation during hypo- or hyperthyroidism respectively), one has to assume that genetical factors take part in the determination of the character of coloration.

Thus, the regional coloration of the plumage is determined by the interference of the hormonal action, on the one hand, and by the differential threshold, on the other. Under typical conditions, a certain level of hormone formation, common for the whole organism, is sufficient to reach the threshold of the reaction in some parts, but is incapable of producing the same effect in other parts, owing to their lesser sensitivity. An increase or a decrease of the activity of the inductor (hormone of the thyroid gland) produces respectively either the reaching of the threshold of reaction on parts previously unsusceptible to the stimulus, or, on the contrary, a total absence of reaction in the most sensitive parts. In either case the result is similar with respect to the disappearance of the regional coloration and different with

respect to the type of the generalized colouring. In the former case (increase of the activity of the inductor) all parts are levelled to those which possess the highest threshold, in the latter (decrease of activity) all parts are reduced to the reaction with the lowest threshold.

The significance of the above hypothesis lies in the generalization of the experimental data already obtained, while it opens the way to further investigation of the problem of the regional coloration of animals.

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PALEONTOLOGY

OCCURRENCE OF GENUS *SCHAUBEUMYS* WOOD (*CRICETIDAE*, *MAMMALIA*) IN THE UPPER OLIGOCENE OF MIDDLE ASIA

By A. I. ARGYROPULO

(Communicated by A. A. Borissiak, Member of the Academy, 7. III. 1939)

Schaubeumys Wood 1935 (1) is a typical representative of fossil *Cricetodontidae*, i. e. of rather primitive hamster-like rodents in the modern concept of this group. By a number of characters it stands close to *Eumys* Leidy. This genus has been hitherto known but by a series of lower molars and one upper molar of the single species *Sch. grangeri* Wood (1) from the Miocene (Pine Ridge Indian Reservation) of South Dakota.

At the present two representatives of *Schaubeumys* have been identified by me among the material from Asia, available in the Paleontological Institute of the Academy of Sciences, but, unfortunately, these are likewise represented by lower molars and fragments of dentale alone.

The following are generic characters of *Schaubeumys*: metaconid on M_1 and M_1 separated from the rest of the cusps of the crown by a rather deep groove. Cingulum anterior labiale enlarged, in size about equal to metaconid, i. e. forming an accessory cusp within the detached first loop. Protoconid with well developed posterior arm (Hinterarm). Cingulum posterior very large, on M_1 and M_1 surpassing the hypoconid. M_3 shorter than in other molar series, bearing in addition to the posterior arm of the protoconid yet a well developed mesostylid spur.

By the general dimensions of the molars and by their patterns, as well as by the absence of a mesostylid spur on M_1 and M_2, *Schaubeumys* is rather close to *Eumys*. The two genera differ by the metaconid in *Schaubeumys* being fused with a greatly enlarged cingulum anterior labiale and separated (on molars unworn or little worn) from the protoconid by a deep groove. It should be noted that in *Eumys*, too, a tendency toward separation of the metaconid on M_1 is observable. Furthermore, *Schaubeumys* is distinguished by the total lack of a pseudo-spur of the mesostylid (Pseudomesostylidsporn) which is always present in *Eumys*.

Schaubeumys aralensis sp. n. № 210.—1. Fragment of dentale with M_1 and M_3 (dextr.). Preservation of bone fair, of teeth excellent. Kept at the Paleontological Institute of the Academy of Sciences of the USSR.

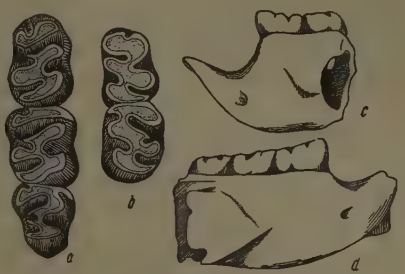
Horizon. Upper Oligocene, Agyspe on the northern shore of the Perovsky bay, Aral Sea, Kazakhstan. Remains enclosed in green clay underlying clay with marly concretions (J. Orlov, 1938).

Length of M_1—M_3 (measured along the crowns)—6.3 mm. The longest and broadest of them (1.95 mm) is the second molar, next in size is M_1, the

third molar being somewhat shorter than the latter, though appreciably narrower, especially in the posterior region. In *Sch. grangeri* Wood the total length of the lower series of molars is 4.38 mm*. The posterior region of M_3 in this genus is not very narrow, being just slightly narrower than the anterior region.

M_1. Cingulum anterior labiale equal in size to the metaconid, fused with it posteriorly, and separated anteriorly by a deep diastema. From the metaconid anteriorwards the border of the crown lacks the small tubercle present in *Sch. grangeri*. The protoconid is large with posterior arm well developed, but not elongate. The ectoconid is large. Hypoconid is small with very strongly developed cingulum posterior, the size of the latter surpassing that of the hypoconid.

M_2. Diastema between metaconid and cingulum anterior labiale very deep, rounded, somewhat displaced upon the lingual side of the tooth. The arrangement of the other elements on M_2 is like that on M_1. On M_3 metaconid fuses with protoconid through a very broad cingulum anterior labiale. The posterior arm of the protoconid equals in size the well developed mesostylid spur. Cingulum posterior is considerably smaller than that of M_1 and M_2. All the characters of dentition indicated above amplify and supplement the generic diagnosis of *Schaubeumys*.



Lower series of molars: *a—Schaubeumys aralensis* sp. n., *b—Sch. woodi* sp. n. Maxillae: *c—Sch. woodi* sp. n., *d—Sch. aralensis* sp. n.

The described genus is to be distinguished from *Sch. grangeri* by the following characters: 1) posterior arm of protoconid shorter and not reaching the edge of the crown; 2) protoconid of M_1 considerably larger; 3) M_1 lacking the accessory cusp in front of the metaconid; 4) the mesostylid heel and posterior arm of protoconid on M_3 much less developed; 5) different sizes and patterns of molars (see above). Outline of lower jaw is shown in the figure.

Schaubeumys woodi sp. n. № 210.—2. Fragment of dentale with M_1—M_2 (sin.). Bone well preserved, teeth in excellent state of preservation. Kept at the Paleontological Institute of the Academy of Sciences of the USSR.

Horizon. See above, description of *Sch. aralensis*.

Surface of molars more worn than in *Sch. aralensis*, up to disappearance of the groove between metaconid and paraconid. Size somewhat inferior to that of the preceding genus (M_1+M_2=4.2 mm upon the crowns). M_1 relatively somewhat narrower and shorter. Position of elements in the crowns of the available molars like that in *Sch. aralensis*. There are essential differences in the shape of the lower jaw. In *Sch. woodi* the diastemal incisura of mandibulae is much deeper than in *Sch. aralensis*, and the foramen mentale is placed below the alveolar border of M_1, the distance being about double the height of the crown of tooth. In *Sch. aralensis* the corresponding distance is approximately that of the height of the crown (figure).

The occurrence of an American mid-tropical genus *Schaubeumys* Wood in the Oligocene of Middle Asia furnishes additional evidence in support of the very close kinship that had once existed between the faunae of America and Asia. After the works by Matthew and Granger(3), who had described a number of rodents from the Tertiary deposits of Mongolia, the community

* In the work by Wood the length of the tooth series is not indicated. The measurements quoted above have been computed upon measurement of the figure in his paper magnified tenfold.

of distribution of a number of genera has been reported more than once. Thus, *Eumys* Leidy, which had been rather abundant in America, was recognized in the Oligocene of Mongolia (*E. Asiaticus* M. et Gr.). The Oligocene genus *Desmatolagus* M. et Gr., known as one of the four species from Mongolia (Ordos and Loh), was identified by Burke (4) as having existed in the Oligocene of Montana, USA. The same author has referred to the Lower Oligocene of Montana the genus *Andynomys* M. et Gr. described from the basin of Tzagan-nora in Mongolia. The genus *Prosciurus* Matthew, widespread and abundant in the Oligocene of America, is represended in Mongolia by one species—*Pr. lohiculus* M. et Gr. Later on a representative of *Prosciurus*, possibly, identical with *Pr. lohiculus* (only the lower jaw with incomplete row of molars is known), has been allocated by us (in litt.), according to the specimen from the collection of the Paleontological Institute of the Academy of Sciences of the USSR, to the far west from the first find in Central Kazakhstan (Oligocene of Myn-sai). On the other hand, Wood (5) inclines to the opinion that an identical species had lived in the Lower Oligocene of Saskatchewan (Cypress Hills).

Like *Prosciurus*, numerous genera of rodents had a very wide range of distribution in Asia. For instance, *Cricetops* M. et Gr. described from the Oligocene of Mongolia (Hsanda-gol formation) was afterwards found in the upper Oligocene of Central Kazakhstan, where it is represented by a very closely allied species (*Cr. affinis* A. Arg.) (2).

Passing to the contemporaneous fauna of rodents in Asia, we are still finding among them traces of ancient community with the fauna of North America; however, such relicts are rather scarce in Asia at the present. In this respect of greatest interest are two genera of rodents *Zapus* and *Calomyscus*, having a great number of allied recent forms only in America. Both of them are represented in Asia by isolated species. *Zapus setshuanus* Pousarg is dwelling in south-western China, *Calomyscus bailwardi* Thos. has a somewhat wider range of distribution in the deserts of Asia Minor.

The study of mammals recovered together with *Schaubeumys* in the Aral plain is of great interest. Alongside of this small hamster, was found a giant hamster-like rodent *Aralomys gigas* gen. et sp. n. (Arg. in litt.) surpassing in size even *Cricetops* (length of its lower row of molars measuring over 11 mm), and a couple of more diminutive hamsters, apparently also representatives of the new genus. Curiously enough, these rodents, like all *Hesperomyidae* [after S. Schaub(6)], exhibit not only a reduced mesostylid spur, but a posterior arm of the protoconid either vestigial or lacking altogether. Furthermore, along with *Schaubeumys* was recovered a primitive representative of *Ochotonidae*, which had still retained roots on the upper molars (*Agispelagus simplex* g. et sp. n. Arg. in litt.), two species of diminutive *Castoridae*, tentatively determined as *Steneofiber*, and species and genus *Dipodidae* subfam. *Allactaginae*, still lacking definition.

In concluding this brief communication, thanks are due to Prof. B. S. Vinogradov and Dr. S. Schaub (Basel) for their valuable advices and literature they kindly placed at my disposal.

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PALEONTOLOGY

RECOVERY OF A REPRESENTATIVE OF *MURINAE* (*GLIRES*, *MAMMALIA*) IN TERTIARY DEPOSITS OF THE USSR

By A. I. ARGYROPULO and I. G. PIDOPLICHKA

(Communicated by A. A. Borissiak, Member of the Academy, 4. III. 1939)

Unlike the family *Cricetidae*, of which numerous forms have been known beginning from the Lower Oligocene, representatives of the family *Muridae* have been found in Tertiary deposits only since the Lower Pliocene (Pontian stage). There are not many finds of fossil *Muridae* known, and such are mainly reported from Western and Southern Europe (from deposits of Monte Bamboli, Pikermi, Samos, Perpignan, Polgardi and some others); whereas isolated finds have been referred ([1],[2]) to Mongolia (Ertemte). As far as finds of *Muridae* from Quaternary deposits are concerned, such are more numerous and known also within the boundaries of the USSR. For the most part, these remains belong to representatives of the recent genus *Apodemus*, Cairo, to the genus *Rhagamys F. Major*, occurring but in the Quaternary deposits of the islands Sardinia and Corsica, and finally, to the genus *Parapodemus* Schaub, known from the Pliocene deposits as well.

Muridae, which had lived toward the end of the Tertiary period, are referred to four genera: *Parapodemus* Sch. (1938), *Progonomys* Sch. (1938), *Anthracomys* Sch. (1938) and *Stephanomys* Sch. (1938). The latter genus is the most peculiar of all, whereas the rest of them approach very closely the recent palearctic genus *Apodemus* and some others. The most characteristic feature of the entire group is the fusion of the elements of the last two loops on M^1 and M^2 (viz., cingulum mediale, protocone, paracone, metacone, pseudo-talon and cingulum posterior linguale), resulting in the formation of a characteristic figure in the form of a crown or rather of a closed curve made up of 5—6 lobes.

The remains of *Murinae* as herein described from the Pliocene of the city of Odessa belong to the genus *Parapodemus* Sch. This genus had a wide range of distribution at the end of the Tertiary period. Thus, in Europe it was found in the Pliocene deposits of Pikermi, Samos, Polgardi, Villard-de-Domsure, Mollon, the age of deposits of the former three localities being identified as Lower Pliocene (Pontian stage). Likewise some representatives of *Murinae* from the Pliocene of Mongolia [«*Mus*» *hipparionum* Schlosser ([1])] are tentatively assigned by S. Schaub (1938) to that genus. Finally, the same author has reported *Parapodemus* from the Quaternary deposits of Europe (Rumania) as well. By its taxonomic peculiarities *Parapodemus* is closely approaching the recent genus *Apodemus*, while differing from it by complete

reduction of cingulum posterior linguale. Being a stable character, the latter may serve to differentiate a whole group of species of Pliocene *Murinae*.

Parapodemus similis sp. n. Mandibulae dextrae № 6585-6587, two of them with complete rows of molars, mandib. sinistrae № 6582-6584, one of them bearing M_1 and M_2, a fragment of maxilla sin. № 6581 with $M^1 + M^2$ and complete plate of ramus inferior processus zygomat. oss. maxillaris.

The bones are mineralized, of reddish and red-brownish colour, preservation—fair. The enumerated specimens belong to the collection of the Odessa Paleontological Expedition of the Academy of Sciences of the Ukrainian SSR, catalogue numbers of the same Expedition. Collected by T. G. Gritzai.

H o r i z o n. Red-brownish clay, filling cracks and funnels in the Pontian limestone, referable to the Middle Pliocene. This clay contains, besides numerous bones of large and small vertebrates (ostrich, camel, mastodon, machairodus, hyaena, bear, fox, «Piping Hare», mole-rat, diminutive hamsters, and so on), so-called Karpathian pebbles.

D e s c r i p t i o n. This species is closely related to *P. coronensis* Sch. (1938) from the Lower Quaternary deposits of Rumania (Fortyogoberg, near Brasso), from which it is distinguished by different pattern of M^1 and details of the arrangement of the elements in the crowns of M^1 and M^2. According to size, *P. similis* sp. n., together with *P. coronensis* and *P. lugdunensis* Sch., are the most minute representatives of the genus. Thus, the dimensions of teeth of the lower jaw № 6585 and of the upper one № 6581 (types!) in *P. similis* are as follows:

M_1	length	1.7,	breadth	1.1	M^1 sup.	length	1.8,	breadth 1.4
M_2	»	1.2	»	1.2	M^2 »	»	1.2	» 1.3
M_3	»	0.9	»	1.0				

$M_1 - M_3 = 3.5 - 3.8$ mm

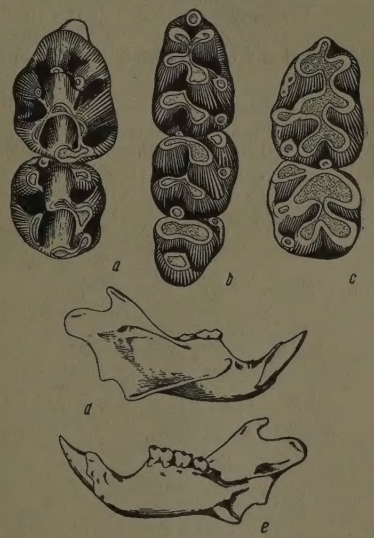
$M^1 + M^2 = 3$ mm

In *P. coronensis* Sch. the length of $M^1 + M^2 = 2.6$ mm, in *P. gaudryi* Dames—3.5 mm. The length of the complete row of the lower series in *P. gaudryi* and *P. vireti* Sch. is from 4.3 to 5.4 mm. The condylar length of the lower jaw in *P. similis* sp. n. is 13.6 mm (±), the height behind M_3—3.5 mm, the dorso-ventral and medio-lateral diameter of the incisor is 1.2×0.6 mm.

M^1—all of the three tubercles of the first loop are disjoined, cingulum anterior linguale being the one most shifted posterad. The valley separating this tubercle from tb 2 is so deep, as to prevent its fusion with tb 2, even though the grinding surface of the molars be so much worn off as to level it down to one plane. Much earlier it is united by a small branch to the protocone*. The above-mentioned character not only serves to sharply distinguish *P. similis* from *P. coronensis* and other species of *Parapodemus*, but is generally not characteristic of *Parapodemus*, being distinctive of the genus *Progonomys* (however, the latter genus possesses, apart from that, a number of other peculiar features). Tb 2 is connected by a narrow branch with cingulum anterior labiale. Notwithstanding the relatively young age of the described specimen (№ 6581) (figure), all of the five cusps of the second half of the tooth have already merged together, forming a crown-shaped figure, so peculiar to genera of the *Apodemus* group. The protocone and pseudo-talon are the largest of them. The described species lacks entirely the tubercle situated below the metacone; such tubercle is well developed, for instance, in *P. gaudryi*, where it is partly fused with the pseudo-talon (cf. fig. Table 1, S. Schaub). Quite characteristic of the described species is the general outline of M^1 (sup.) which is relatively very broad and short (figure). The diastema on the border of the crown, between tb 1 and 2, is relatively very weak.

* Nomenclature of the molar elements mainly after S. Schaub.

M^2—the two anterior cusps (cingulum anterior linguale and labiale) are very well developed, a large c. ant. linguale is directed a little way back from the anterior border of the crown, sending off a branch toward the paracone. Another peculiarity is the presence of a small tubercle placed at the very edge anteriorwards from c. anterior linguale, as if displacing the last tubercle backwards. Of the main cusps, the paracone and protocone are joined to one another by a fine branch, while the elongated cingulum mediale (tb 4) lies more isolated than on the first molar, and carries a vestigial branch. Metacone and pseudo-hypocone are confluent; the metacone lying nearly at the very edge of the crown, and rather suggestive of the small accessory cusp at the edge of the crown in *P. gaudryi* (cf. fig. Table 1 in S. Schaub's work), is very characteristic of the described species. This is apparently conditioned by a very short labial branch of the metacone in the described species. A similar position of the metacone is seen in *Anthracomys majori* Sch., the latter, however, displaying a number of other differential features.



Parapodemus similis sp. n.: *a* (№ 6581)—M_1—M_2 sin.; *b* (№ 6585)—M_1—M_3 dext.; *c* (№ 6584)—M_1—M_2 sin.; *d*, *e*—mandibulae (№ 6585), outer and inner view.

M_1—the anterior accessory cusp is very well developed, the pair of innominate cusps symmetrically disposed, while the paraconid is shifted very strongly backwards from the metaconid, as well as the hypoconid. There are two cuspules on the labial border of the crown (in all three specimens *P. similis*) both arising near the paraconid and hypoconid, and fusing with it on the more worn molars. Orad from the anterior cusp two more very small and indistinct ones could be placed.

Such cusps are present in 2 specimens. The terminal cusp of the crown, which should evidently be considered as a modification of the hypoconid branch or cingulum posterior (3), is weakly developed. The same applies to M_2.

M_2—paraconid and hypoconid are also displaced backwards, but less so than on M_1. In addition to cingulum anterior labiale, it bears another tubercle arising at about the middle of the labial border, near the paraconid.

M_3—the pattern of this molar is one usually exhibited in *Apodemus* or *Parapodemus*. Its peculiar feature is the presence of a small tubercle (cingulum anterior labiale), placed on a high base and separated by a narrow, deep valley from the paired cusps of the first loop.

It follows from the above description that *P. similis* sp. n. is distinguished from all the known species of *Parapodemus* by a very deep valley between cingulum anterior linguale and tb 2 on M^1, preventing the fusion of those tubercles, even in very old specimens. As to less significant differentia, *P. similis* sp. n. differs from species coming close to it by their size in that M^1 of *P. similis* has a sharply pronounced bunoid form of the crown, its width to length index being 0.777, whereas in *P. coronensis* this ratio equals 0.650; the same character marks off *P. similis* from *P. lugdunensis* Sch. (1938). Besides, *P. similis* displays on M^2 a cingulum anterior labiale of greater size than in *P. coronensis*, and M^1 in *P. similis* lacks the accessory

cuspule on the posterior border of the crown behind the metacone, this cuspule being manifest in *P. lugdunensis*.

The lower jaw in *P. similis* (figure) differs sharply from the jaws of recent representatives of *Apodemus* by relative robustness, a high median portion of dentale and deep diastemal incisura. By relative shortening and depth of the diastemal incisura the jaw in *P. similis* recalls that of *Micromys minutus* Pall.

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